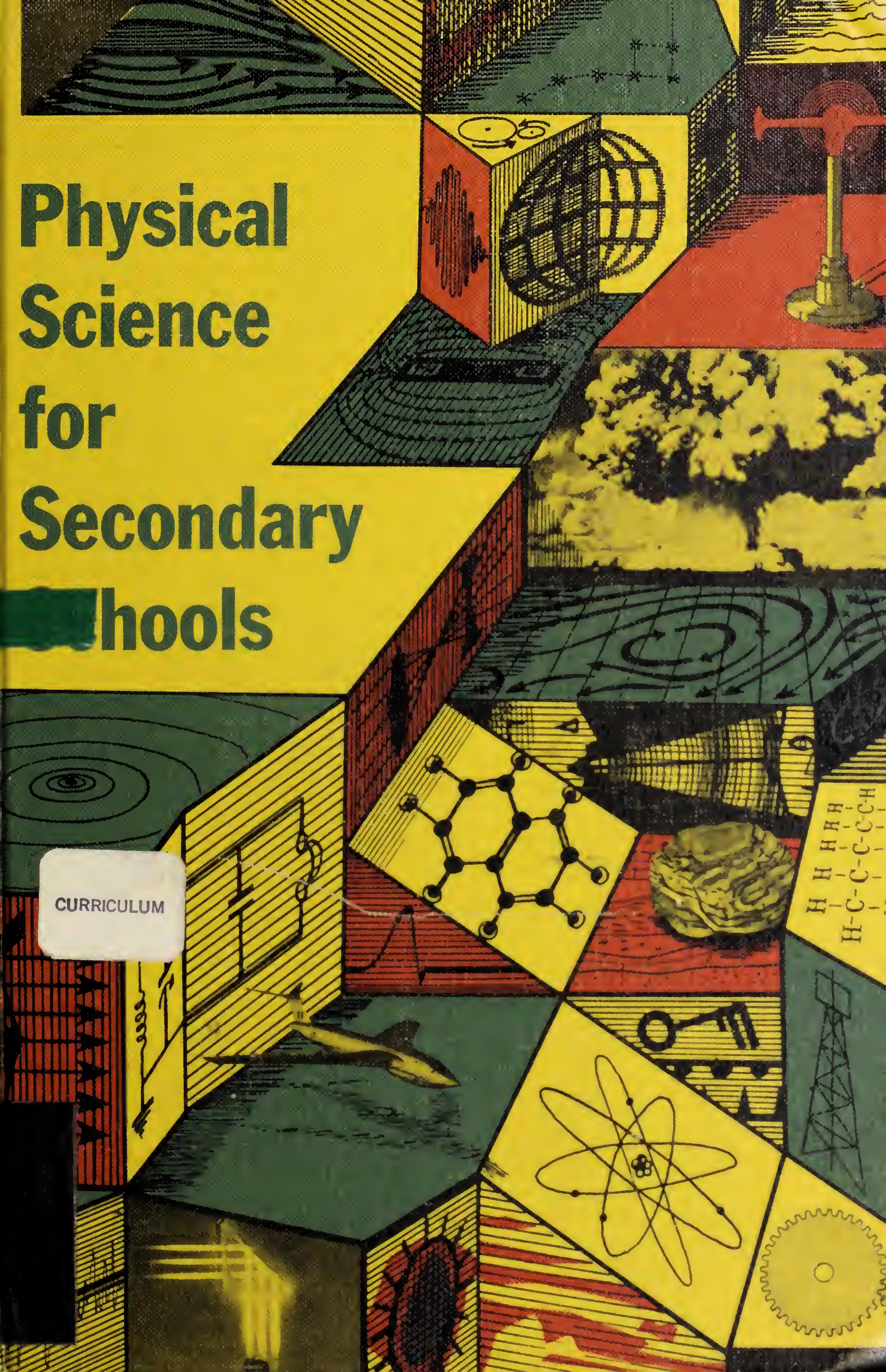
Physical Science for Secondary hools



CURRICULUM

THE ELEMENTS

ELEMENT	SYMBOL	ATOMIC NUMBER	ATOMIC WEIGHT
Actinium	Ac	89	227
Aluminum	Al	13	26·98
Americium	Am	95	(243)
Antimony	Sb	51	121·76
Argon	A	18	39·944
Arsenic	As	33	74·91
Astatine	At	85	(211)
Barium	Ba	56	137·36
Berkelium	Bk	97	(245)
Beryllium	Be	4	9·013
Bismuth	Bi	83	209·00
Boron	B	5	10·82
Bromine	Br	35	79·916
Cadmium	Cd	48	112·41
Calcium	Ca	20	40·08
Californium	Cf	98	(248)
Carbon	C	6	12·011
Cerium*	Ce	58	140·13
Cesium	Cs	55	132·91
Chlorine	Cl	17	35·457
Chromium	Cr	24	52·01
Cobalt	Co	27	58·94
Copper	Cu	29	63·54
Curium	Cm	96	(245)
Dysprosium*	Dy	66	162·51
Einsteinium	E	99	(255)
Erbium*	Er	68	167·27
Europium*	Eu	63	152·0
Fermium	Fm	100	(252)
Fluorine	F	9	19·00
Francium	Fr	87	(223)
Gadolinium*	Gd	64	157·26
Gallium	Ga	31	69·72
Germanium	Ge	32	72·60
Gold	Au	79	197·0
Hafnium	Hf	72	178·58
Helium	He	2	4·003
Holmium*	Ho	67	164·94
Hydrogen	H	1	1·0080
Indium	In	49	114·82
Iodine	I	53	126·91
Iridium	Ir	77	192·2
Iron	Fe	26	55·85
Krypton	Kr	36	83·8
Lanthanum*	La	57	138·92
Lead	Pb	82	207·21
Lithium	Li	3	6·940
Lutecium*	Lu	71	174·99
Magnesium	Mg	12	24·32
Manganese	Mn	25	54·94
Mendelevium	Mv	101	(256)
Mercury	Hg	80	200·61
Molybdenum	Mo	42	95·95
Neodymium*	Nd	60	144·27
Neon	Ne	10	20·183
Neptunium	Np	93	(237)
Nickel	Ni	28	58·71
Niobium	Nb	41	92·91
Nitrogen	N	7	14·008
Nobelium	No	102	(253)
Osmium	Os	76	190·2
Oxygen	O	8	16·0000
Palladium	Pd	46	106·7
Phosphorus	P	15	30·975
Platinum	Pt	78	195·09
Plutonium	Pu	94	(242)
Polonium	Po	84	210
Potassium	K	19	39·100
Praseodymium*	Pr	59	140·92
Promethium	Pm	61	(145)
Protactinium	Pa	91	231
Radium	Ra	88	226·05
Radon	Rn	86	222
Rhenium	Re	75	186·22
Rhodium	Rh	45	102·91
Rubidium	Rb	37	85·48
Ruthenium	Ru	44	101·1
Samarium*	Sm	62	150·35
Scandium	Sc	21	44·96
Selenium	Se	34	78·96
Silicon	Si	14	28·09
Silver	Ag	47	107·880
Sodium	Na	11	22·991
Strontium	Sr	38	87·63
Sulphur	S	16	32·066
Tantalum	Ta	73	180·95
Technetium	Tc	43	(99)
Tellurium	Te	52	127·61
Terbium*	Tb	65	158·93
Thallium	Tl	81	204·39
Thorium	Th	90	232·05
Thulium*	Tm	69	168·94
Tin	Sn	50	118·70
Titanium	Ti	22	47·90
Tungsten	W	74	183·86
Uranium	U	92	238·07
Vanadium	V	23	50·95
Xenon	Xe	54	131·30
Ytterbium*	Yb	70	173·04
Yttrium	Y	39	88·92
Zinc	Zn	30	65·38
Zirconium	Zr	40	91·22

*Rare Earths

Atomic weight in brackets denotes the mass number of the isotope of longest known half-life.

REFERENCE TABLES

Table 1. *Useful Values and Relationships*

12 in. = 1 ft.
3 ft. = 1 yd.
5280 ft. = 1 mi.
1760 yds. = 1 mi.
6080 ft. = 1 nautical mile
144 in.^{2} = 1 ft.^{2}
9 ft.^{2} = 1 yd.^{2}
1728 in.^{3} = 1 ft.^{3}
27 ft.^{3} = 1 yd.^{3}
231 in.^{3} = 1 gal.
60 mi/hr. = 88 ft/sec.
70000 grains = 1 lb.
16 oz. = 1 lb.
2000 lb. = 1 ton

1 ft.^{3} of water weighs 62.4 lb. wt.
1 liter of air at S.T.P. weighs 1.29 gm. wt.
1 B.t.u. = 252 calories
Speed of sound = 1090 ft/sec. at 0°C.
Circumference of a circle = $2\pi r$
Area of circle = πr^2
Surface of a sphere = $4\pi r^2$
Volume of a sphere = $4/3\pi r^3$
Standard temperature = 0°C
Standard pressure = 760 mm. of mercury
Latent heat of vaporization of water 540 cal/gm. or 972 B.t.u/lb.
Latent heat of fusion of ice 80 cal/gm. or 144 B.t.u/lb.

Table 2. *Measurements in the Metric System*

LENGTH

1000 meters = 1 kilometer (km.)
100 meters = 1 hectometer (hm.)
10 meters = 1 dekameter (dkm.)
0.1 meter = 1 decimeter (dm.)
0.01 meter = 1 centimeter (cm.)
0.001 meter = 1 millimeter (mm.)
0.000001 meter = 1 micron (μ)

WEIGHT

1000 grams = 1 kilogram (kg.)
0.001 gram = 1 milligram (mg.)

AREA

Area is expressed as square millimeters (mm.^{2}) or square centimeters (cm.^{2}), etc.

VOLUME

Volume is expressed as cubic millimeters (mm.^{3}) or cubic centimeters (cm.^{3}), etc., *or* as liters (1.) (1 liter is the volume of 1 kilogram of water at 4°C) or milliliters (ml.) (1 milliliter is approximately 1 cubic centimeter).

Table 3. *Metric-English Equivalents*

1 in. = 2.5400 cm.
1 ft. = 30.480 cm.
1 yd. = 91.440 cm.
1 mi. = 1609.4 m.
1 mi. = 1.6094 km.
1 mm. = 0.03937 in.
1 cm. = 0.3937 in.
1 m. = 39.37 in.
1 m. = 3.2808 ft.
1 m. = 1.0936 yd.
1 km. = 0.621 miles
1 in.^{2} = 6.4516 cm.^{2}

1 cm.^{2} = 0.1550 in.^{2}
1 in.^{3} = 16.3872 cm.^{3}
1 cm.^{3} = 0.0610 in.^{3}
1 grain = 0.06480 gm.
1 oz. = 28.3495 gm.
1 lb. = 453.592 gm.
1 lb. = 0.4536 kg.
1 gm. = 15.4324 grains
1 gm. = 0.03527 oz.
1 gm. = 0.002205 lb.
1 kg. = 2.2046 lb.
1 liter = 0.88 (imperial) qt.

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Physical Science for Secondary Schools

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PREFACE

This book, PHYSICAL SCIENCE FOR SECONDARY SCHOOLS, is part of a basic course which should provide interesting reading even for students who, hitherto, have had little or no contact with science as a formal study.

It is intended that the book should be used to follow a course in physical science that has covered such subjects as machines, heat, and light. This book therefore follows logically after BASIC PHYSICAL SCIENCE by Hogg and Cross.

The text is organized into seven units, subdivided into thirty-six related chapters. At the end of each chapter are "Things to Remember," "Questions," and "Things to Do." The questions are arranged, according to difficulty, into A and B groups, the latter being more exacting.

Throughout the text there are directions for numerous demonstrations. These should be performed, not merely discussed, since students like to see things happen. If the demonstrations are well performed, students will be stimulated to do the projects which follow each chapter and even to devise projects of their own. They should be required to do most of the suggested projects as part of their laboratory assignments.

The authors wish to acknowledge the help and assistance that have been given by the following people: Professor R. W. Longley and Professor C. R. Stelck, University of Alberta, who reviewed the sections on weather and geology respectively; Mr. O. Massing, Department of Education, Province of Alberta; Mr. F. M. Riddle, Superintendent of Schools; Mr. N. J. Andruski, Inspector of High Schools; Professor D. B. Scott, University of Alberta; Mr. E. N. Simbalist, Mr. S. W. Radomsky, Mr. N. J. Pickard, and Mr. A. Russell, science teachers.

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Physical Science for Secondary Schools

The Earth As It Began

LOOKING AHEAD

This unit raises some ot the ultimate questions that have puzzled philosophers ever since man began to think. They concern "the beginning." How was the earth formed? How was the earth divided into continents and seas? How did the earth acquire its atmosphere? Is the earth the only planet in the universe that can support life? How were the stars formed? And the Universe?

Even today, after pondering these questions for centuries, neither philosophers nor scientists would claim they know the answers. During the past 50 years more scientific knowledge has been acquired than in all previous centuries. Scientists have applied the new knowledge in seeking answers to these cosmic problems. The answers, however, are far from complete, and many gaps in our knowledge have been bridged by intelligent speculation. But this is as it should be; this is the way knowledge advances. The search will continue. Astrophysicists will discover new knowledge and slowly, through the years—maybe through the centuries—the speculations of today will be confirmed or rejected.

←—— *As the earth cooled the rains came, and then vegetation. And the bleak earth lost some of its desolation.* (Ewing Galloway)

1

THE EARTH— ITS ORIGIN

When we look at the sky on a clear autumn night we are impressed by the immense size of the universe. With the naked eye we can see four or five thousand stars. If we look through binoculars we can see almost a hundred thousand, and if we scan the heavens with a large telescope we can see millions of them. What are stars? How are they formed?

Ever since the beginning of recorded history man has looked at the heavens with awe and wonder, and has speculated upon the origin of this vast panorama which is ever before him. Its mysteries were the concern of all the ancient religions and philosophies. You may recall that the writer of the Book of Genesis begins his first chapter with the inspiring and resounding phrases, "In the beginning God created the heaven and the earth. And the earth was without form, and void; and darkness was upon the face of the deep. . . ."

From earliest times men of religious insight have pondered the deep mysteries of the universe, and the Bible asks many questions about them. Nowhere are they more beautifully expressed than in the 38th chapter of Job, which begins, "Where wast thou when I laid the foundations of the earth? Declare, if thou hast understanding." With the growth of knowledge the scientist has answered some of these ancient questions, but others remain unanswered.

Scientists confine themselves to things that can be measured and recorded; they deal with the physical universe and the various aspects of life on earth. They have learned a great deal by searching the heavens with their giant telescopes, and are beginning to understand what is happening in the outermost parts of the visible universe. Their store of knowledge is ever increasing, but no sooner is one question answered than others suggest themselves. This is the inevitable consequence of man's inquiring mind. We shall discuss some of the scientist's ideas concerning the creation of the earth. But we must bear in mind that even the scientist speculates in this field where his knowledge is far from complete. It is, therefore, not surprising that even among scientists there are differences of opinion.

Earth, A Planet of the Sun. Throughout most of the long span of human history it was generally believed that the earth was the center of the universe. Then, barely 400 years ago, Copernicus, a Polish astronomer, proved that the earth revolves about the sun and that it is merely one of the planets in the solar system. But what of the sun? Is it the

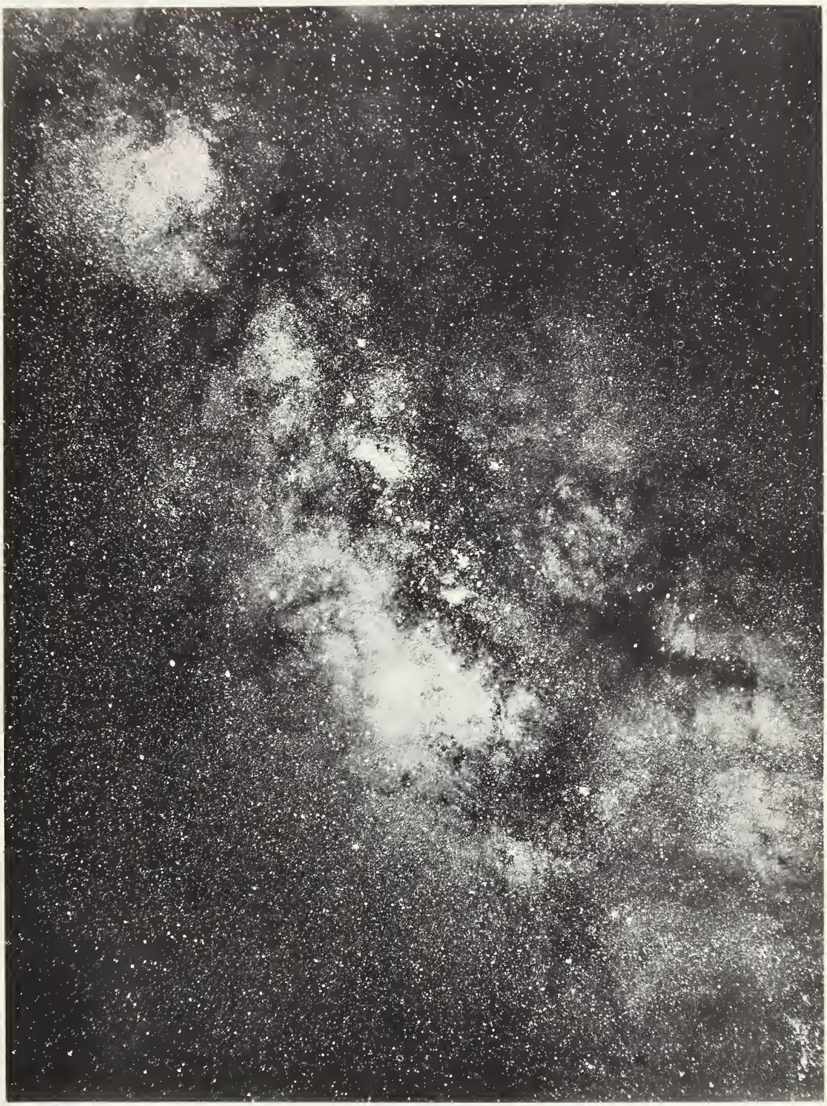


Fig. 1-1. View towards the center of the Milky Way where the concentration of stars is greatest. All stars seen by the unaided eye and most of those seen through the most powerful telescopes are in the Milky Way.

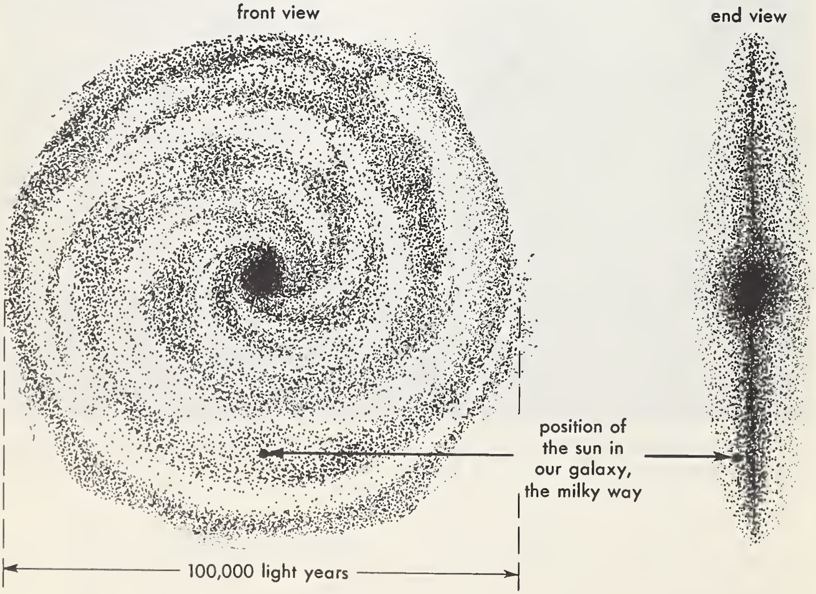
center of the universe? As we gaze into the heavens with a telescope, the motions of the nine planets can easily be followed, but the positions of the stars appear to be fixed. They still seem to be in the positions they occupied when they were described by the ancient Greek astronomers 2000 years ago. How can this be explained? Again by looking through a telescope, we learn that the sun is merely a member of an enormous system of stars which we call a galaxy. And the name given by the ancient Greeks to this particular galaxy is the Milky Way.

The Milky Way. The Milky Way is an enormous agglomeration of a hundred billion stars. In shape it is like a wheel, or lens-shaped if viewed end on. It is so vast that light, which travels at 186,000 miles per second, takes 100,000 years to cross from edge to edge. For this reason, astronomers call the distance through the diameter of the Milky Way 100,000 light-years. (A light-year is the distance that light travels in one year.) Our solar system occupies a position about half-way from the center to the edge of the galaxy, and the thickness of the Milky Way in this region is about 10,000 light-years. Actually, the Milky Way as we see it, is a view from inside the galaxy in which we live. That is why the stars of the Milky Way seem to spread out like a great veil across the sky.

Does the Milky Way move? Although the stars seem to occupy the same positions through centuries of time yet, from their observations, astronomers tell us that the galaxy rotates about its centre. Moreover, they have computed that a complete rotation of the galaxy takes one hundred million years. Stars therefore do not remain in fixed positions; their positions appear to be fixed only because of the enormous distances between the earth and the stars.

The Universe. Does the Milky Way constitute our known universe? Besides the stars of the Milky Way, the telescope reveals dark patches of matter which ob-

Fig. 1-2. The front view of the Milky Way galaxy is wheel-shaped; the end view is lens-shaped.



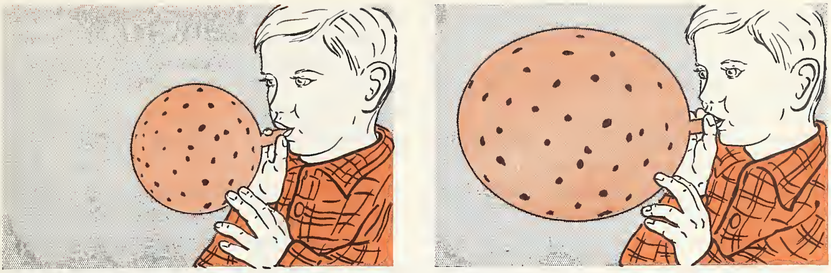


Fig. 1-3. The black spots on the inflated balloon represent galaxies in an expanding universe.

scure the light from some stars, and also dim blobs of luminous matter. These dim blobs have now been identified as other galaxies, far beyond our Milky Way. The actual number of galaxies in the universe is not known but, from photographs already taken by the 200-inch Palomar telescope, it is estimated that there are, within the range of this great telescope almost a billion (1,000,000,000) of them. And probably there are many more beyond its range. No one knows.

This, then, is the astronomer's staggering concept of the universe—billions of galaxies each containing billions of stars.

An Expanding Universe. Still another extraordinary concept that astronomers are agreed upon is that all the galaxies are rushing away from each other at enormous speeds. In other words, the universe is expanding. We can get a crude picture of the expanding universe by inflating a balloon whose surface is dotted by black spots to represent the galaxies. As the balloon expands the spots recede from each other.

How do astronomers account for this strange behavior? One theory is that in the beginning all the galaxies were bunched together within a relatively small space. There was nothing but a mass of enormous density and exceedingly high temperature. Then, about six billion years ago, there was an enormous explosion. This was the moment of Creation, for the stars were made out of the dust of the explosion.

How Were the Stars Formed? Most astronomers agree that the stars were made out of the receding clouds of gas formed in that colossal explosion. Under the influence of the gravitational force of attraction, the gas condensed to dust, and the dust particles cohered, contracted, and rotated. The contraction caused the temperature to rise until, eventually, it became white hot and a star evolved.

Our sun, for example, was probably formed in this way from an enormous cloud of cosmic dust. As it was compressed it began to glow and rotate. The distance across this enormous rotating disc was roughly the same as the diameter of the present solar system. As the disc continued to spin, centers were formed within the cloud, about which some of the whirling matter condensed. These whorls formed the planets and, among them, our earth. Thus, according to this theory, our earth came into being as a ball of matter which became molten and white hot as it was squeezed to shape by the force of gravitation.

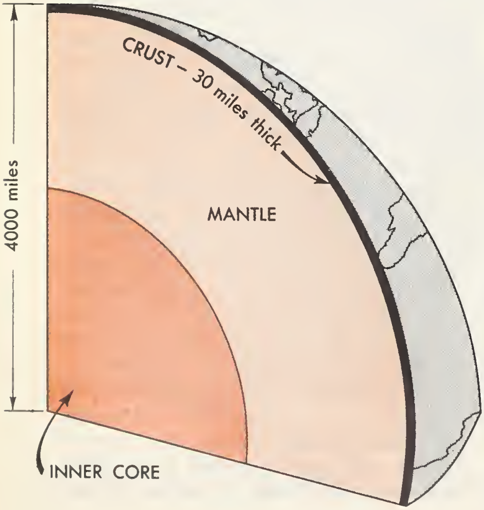
The Formation of Continents. While the earth was still molten, gravitation caused a separation of its constituents: the heavy elements sank towards the center and the lighter ones rose to the surface. Eventually, after centuries of time, sufficient cooling took place on the surface so that solid rock began to crystallize. These first surface rocks were granite (see page 48). As crystallization continued, the granite blocks, floating on a sea of molten rock, grew outward in all directions until they were of enormous size.

These blocks of granite were the first land masses and comprised the original continents. Strange as it may seem, most geologists believe that these original granite continents mark, with little change, the positions of the continents today. The deep valleys or basins between the granite blocks became filled with water in the manner described in Chapter 2. In this way the oceans were formed, and geologists also believe that the positions of the large oceans have remained unchanged throughout the ages.

Through the succeeding centuries the earth's interior cooled and contracted. As a result, the outer crust became too big for the interior and, in settling, the crust became wrinkled just as the skin of an apple becomes wrinkled during baking. But these wrinkles on the earth's surface were mountains. In this way, many times during geologic history mountains have been formed—a topic discussed in Chapter 8.

The Structure of the Earth's Interior. Geologists tell us that the earth consists of three concentric spheres and resembles a gigantic melon, 8000 miles in diameter. The innermost sphere, 4000 miles in diameter, is composed of a heavy molten material, probably iron and nickel. The geologist knows this central part is liquid because it behaves like a liquid in being unable to transmit vibrations of a certain kind set up by earthquake shocks.

Fig. 1-4. The earth consists of three concentric spheres. The inner core is probably molten iron; the outermost shell is a thin crust, 25 to 30 miles thick.



Resting upon the core of molten iron is an inner stony shell, probably of olivine (see page 46), about 2000 miles thick. This stony mantle, which reaches almost to the surface, is a dense plastic rock probably red-hot throughout its whole thickness.

The outermost shell is the thin crust, about 30 miles thick. The crust is made up of two main layers. The lower layer, 20 to 25 miles thick, is a shell of hard black rock called basalt (see page 49). The basalt layer lines the beds of the oceans and forms the underpinning supporting the continents. As already stated, the uppermost layer is made of granite. It is not continuous like the basalt layer, but comprises the continental masses upon which we live.

As we shall read in Chapter 7, some of the uppermost granite layer has been weathered and eroded, so that sedimentary deposits have been laid upon it. Nevertheless, the general picture remains. Progressing downward from the crust to the molten iron core, the materials become more and more dense and the temperature steadily rises.

Are There Other Habitable Earths in the Universe? The above description is the brief story of the origin of our earth, the peculiar planet that supports life. Are there other planets which support life or are we alone in the vast universe? In order to support life, a planet must meet a number of conditions. These are concerned with temperature, water, and air.

Temperature. The temperature must not be too high nor too low, since the temperature range within which animal and vegetable life can flourish is very small. We know, for instance, that the planet Mercury cannot support life. It is much too hot on the sun side and much too cold on the side never exposed to the sun.

Water. Liquid water is necessary for all the life processes. If our earth were nearer the sun the water would be boiled to steam; if it were farther from the sun it would be frozen to ice. In either case, life on earth would be impossible.

Atmosphere. Life, as we know it, needs an atmosphere rich in oxygen. The only planet with such an atmosphere in the solar system is our earth. The atmospheres on Jupiter and Saturn, for example, consist of ammonia and methane, and are poisonous.

In other words, to support life the climate, the atmosphere, and other conditions must be just right. Do these conditions exist elsewhere? Let us recall that there are billions and billions of stars in the universe, and from this vast number it is probable that on some planets the conditions of our earth are duplicated. Indeed, a Harvard astronomer, Dr. Harlow Shapley, has estimated that there may be millions of such planets in the universe that support life. He also points out that animal and vegetable life on other planets may be quite different from their counterparts on earth. They may even have surpassed the terrestrial forms. That is to say, insects such as bees, birds such as hawks, animals such as dogs, and even man may have reached more advanced stages of life on other planets. These are interesting speculations.

Things to Remember

The ***earth*** is a planet in the solar system.

The ***sun*** is a star in the Milky Way system.

The ***Milky Way*** is the galaxy of stars that includes the sun.

The ***Universe*** is made up of billions of galaxies.

Stars were probably formed out of enormous clouds of cosmic dust.

The original ***continents*** of the earth were blocks of granite that crystallized from the molten earth.

The ***earth*** probably consists of (1) an innermost sphere of molten iron, (2) a massive inner sphere of olivine rock, and (3) an outermost thin crust of granite and basalt.

The ***crust*** of the earth is 20 to 30 miles thick.

The earth can support life because it has abundant water, oxygen in its atmosphere and an equable climate.

Questions

GROUP A

1. Who was Copernicus? When did he live? What was his contribution to astronomy?
2. What is (*a*) a galaxy, (*b*) a light-year?
3. What is the age of the universe?
4. What are the three factors that determine whether a planet can or cannot support life?

GROUP B

5. Briefly describe the Milky Way.

6. What is meant by the term, ***an expanding universe?***
7. The sun was probably formed from a cloud of cosmic dust. How?
8. Explain how the first continents of the earth were formed.
9. Briefly describe the composition of the earth's crust.
10. The earth is a ***peculiar*** planet, able to support life. What does this statement mean?

Things to Do

To compare the size of the Milky Way with the size of the universe · With the naked eye we can see a distance of 2 million light-years. All the stars we see are in the Milky Way. The Hale 200-inch telescope on Mount Palomar can see a distance of 2 billion (one thousand million) light-years.

Make a square of stiff paper, the sides of the square being at least two meters long. Draw a base line near the edge of one side of the square. From the center of the base line draw a semicircle of one millimeter radius. From the same center draw another semicircle of one meter (100 centimeters) radius.

The very small circle represents the limit of vision with the naked eye and the billions of stars of the Milky Way galaxy lie within it. The larger circle represents the limit of vision through the most powerful telescope. And within its space are the billions of galaxies, similar to the Milky Way.

The nearest galaxy to the Milky Way (Andromeda), is 2 million light-years away. Mark its position on the paper.

2

WATERS OF THE EARTH

THE EARTH is the only planet with an abundance of water. In fact, water is so abundant that the earth is almost drowned by it. More than two-thirds of the surface of the earth is covered by water, and only one-fifth of the Southern Hemisphere is dry land. Indeed, if the earth were a smooth sphere, that is if its mountains and valleys were flattened out, water would cover the whole earth to a depth of a mile and a half.

Water is, of course, a liquid, and liquids in nature are exceptional; the white hot gases of the stars are the usual form of matter. Only within a narrow range of temperatures can water exist, and only within a fraction of this range can life be supported. Water is essential to all forms of life; all animals need it in abundance. Not only do they need it to keep alive, but they are made up largely of water. A man's body is about 70 per cent water, one-third of it is in the blood and two-thirds in the cells. For example, a man weighing 180 pounds carries 125 pounds of water in his various organs. Vegetables and fruits contain even higher percentages of water. Potatoes, for instance, are about 78 per cent water; apples about 85 per cent. All ordinary foods contain a good deal of water. It makes up about 50 per cent of bread, "the staff of life," and about 87 per cent of milk.

The Origin of the Seas. How did water get on the earth in the first place? Or, in other words, how were the seas created? The writer of the 95th Psalm was pondering this same mystical problem when he wrote "In His hand are the deep places of the earth; the heights of the mountains are His also. The sea is His, and He made it; and His hands formed the dry land."

Most scientists agree that the seas came from the molten interior of the earth, that they are a product of volcanic activity. They believe that, as the rocks of the crust crystallized and hardened, water vapor and carbon dioxide were released and rose to the surface. Through the centuries, as the interior cooled, enough vapor was released to form dense clouds that covered the earth. So thick was this mantle of cloud that the sun could not penetrate it and the earth was shrouded in darkness. In time the atmosphere cooled so that some of the vapor condensed to rain, but only to be changed back to steam as soon as it touched the hot earth. Eventually the surface cooled sufficiently so that the rain water was not boiled away. Then came the rains, and the deluge. It probably rained for centuries, and water collected in the basins or fell on the continental masses from

Fig. 2-1. "In His hand are the deep places of the earth; the heights of the mountains are His also. The sea is His, and He made it. . . ."

which it poured into the depressions. This is probably the way the seas were first formed on the earth.

But the basins were by no means filled in that original deluge. Volcanic activity has continued through the ages and with it more water has been released from the interior until, probably about a billion years ago, the seas reached their present levels.

The Salt of the Sea. The first seas were fresh water. Today they are salty and bitter to taste. As the rain water flowed over the land some of the minerals provided nourishment for the minute organisms in the sea upon which all marine life depends.

The salinity of the sea has gradually increased until it now contains in solution about 3.5 per cent of solid material. If 100 pounds of sea water are evaporated to dryness, there will be a white residue weighing about 3.5 pounds. Most of this residue, as you know, is common salt or, as the chemists say, sodium chloride. There are, however, many substances besides sodium chloride in this residue, but these will be referred to in a later chapter.

We must not conclude that all sea water contains 3.5 per cent of dissolved solids. This is a so-called average value; in some seas the percentage is higher than this, in others lower. If a sea has no access to the oceans the amount of solid in solution is much higher than 3.5 per cent. For example, the water of the Great Salt Lake in Utah contains about 30 per cent of salt. How do you account for the high percentage of salt in this inland sea?

Solutions. Sea water has been referred to as a solution. What is a water solution or, as it is usually called, a ***solution?*** It is simply a mixture of water and the dissolved substance. Let us make a solution that roughly corresponds to sea water. We can do this by dissolving 3.5 ounces of salt in 96.5 ounces (or 6 pounds) of water, (notice that 3.5 oz. of salt plus 96.5 oz. of water give 100 oz. of solution) or 0.58 ounce of salt to one pint of water. How are these figures obtained? (The following may seem complicated, in which case it may be omitted.)

Six pounds of water is roughly six pints, and six pints is too much water for convenience. If we take one pint of water we shall need $\frac{3.5}{6}$ or 0.58 ounces. But the scales in your laboratory may be graduated in grams. Grams are a unit of weight in the metric system, and ***one ounce equals 28.4 grams,*** or 1 oz. = 28.4 g. Hence 0.58 ounce is equal to 0.58 × 28.4 or 16.5 grams.

Demonstration

2-1. To make sea water

Add 0.58 oz. (or 16.5 g.) of salt to 1 lb. (or 1 pint) of water in a large beaker. Stir the water with a glass rod until the salt has completely dissolved. In the clear solution the salt can no longer be seen. If we withdraw a few drops of solution and taste it, we realize it is salty no matter from which part of the beaker we extract the drops. The salt has therefore spread throughout the water. How is this possible? This is possible only because there are spaces between the particles or molecules (see page 21) of water, spaces into which the minute particles from the solid salt can go. That is how a solid dissolves.

To imitate the Great Salt Lake the water would have to be 30 per cent salt. Let us make a 25 per cent solution. (A 30 per cent solution is almost saturated and the salt would take a long time to dissolve.) We must now add 25 ounces of salt to 75 ounces, or 4.7 pounds (or 4.7 pints) of water. Or, better still, to every pint of water we must add $\frac{25}{4.7}$ or 5.3 ounces of salt.

Demonstration

2-2. To make Salt Lake sea water

Add 5.3 oz. (or 150 g.) of salt to the pint of water in a beaker. Stir the water with a glass rod. It is difficult to get this large amount of salt into solution by stirring, and it will be helpful to warm the solution. Even the warm solution needs considerable stirring before all the salt dissolves.

It is evident that there is a limit to the amount of salt that will dissolve in a body of water. Great Salt Lake is almost saturated with salt; that is, this solution contains almost as much as it can hold. Still another important thing to remember is, ***the higher the temperature the more solid will dissolve in water.***

Some solids are more soluble than salt, sugar for example. Other solids are less soluble than salt; chalk, for example, is almost insoluble. Rocks, in general, are not very soluble. But when we think that rain has been falling on the land, and rivers have been flowing along their rocky beds, for millions of years the total amount of rock material that dissolves is enormous. We shall discuss the slow dissolving action of running water, a process called ***erosion,*** at some length in Chapter 7.

Rain. Most of the rain that falls on the land comes from the sea. How can we explain this? The heat of the sun causes sea water to evaporate and the vapor rises into the air just as it does when water is boiled in a kettle. The vapor is then carried away by winds. Across Canada the prevailing winds are westerly; that is, they blow ***from*** the west. In this way, water vapor from the Pacific is carried to the western shores of Canada and, if the vapor is cooled sufficiently, it is condensed to clouds and then to rain. The lush dense forests of British Columbia are due to rains that come from the Pacific. But how do Quebec and the Atlantic Provinces get their rain? Most, of course, comes from the Atlantic Ocean, but some of the vapor originates in the warm waters of the Gulf of Mexico. It is carried by air currents up the wide Mississippi Valley and then eastward by the prevailing winds.

Vast quantities of water vapor are carried by winds to bathe the earth as rain. It has been calculated that, for millions of years, rain has been falling on the land at the rate of about four million gallons per second. This steady downpour has, through the ages, changed the face of the land.

Oceans. Actually there are no separate oceans; they are all parts of a single sea that encircles the earth. The largest of the oceans are the Pacific and Atlantic. The Pacific alone covers an area larger than that of all the continents together.

The ocean boundaries appear to be permanent. At least we get this impression when we make our annual visits to coastal waters. The tidal waters advance and recede day after day, year after year, but the boundary between land and water seems to remain in the same place. Yet there is plenty of geological evidence to prove that this is a false impression. Many times during the ages the seas have invaded the land. At times they have encroached upon the continent of North America; at other times they have receded far from its shores so that the continent has "grown" in size. How is this possible?

It is obvious that the sea level would be lowered if large amounts of water were "locked up" as ice on the land; and that sea level would be raised if large amounts of ice melted to water and ran into the seas. Measurements taken by tide gauges prove that the water on the east and west coasts of Canada is slowly rising. This may be due to the melting of polar ice, although other possible causes will be considered later. Suppose, however, that the ice which covers Greenland and the antarctic were to melt. Computations show that, in such a case, sea level would rise at least 100 feet. How would this affect the eastern coast of North America? Most of its cities would be submerged and the Atlantic would extend to the foothills of the Appalachians. Moreover, the sea would encroach over the low land bordering the Gulf of Mexico and the sea would cover most of the wide Mississippi Valley. This is not a hypothetical case; it has actually happened many times during the long span of geologic time.

The Ocean Basin. For many centuries it was thought that the ocean floor was a flat, featureless plain. However, when soundings were made with a plumb (a lead weight attached to a long hemp rope), it was discovered that the floor is as rugged as dry land. Even deep gorges and high mountains were discovered in this way. But the lead-line method of

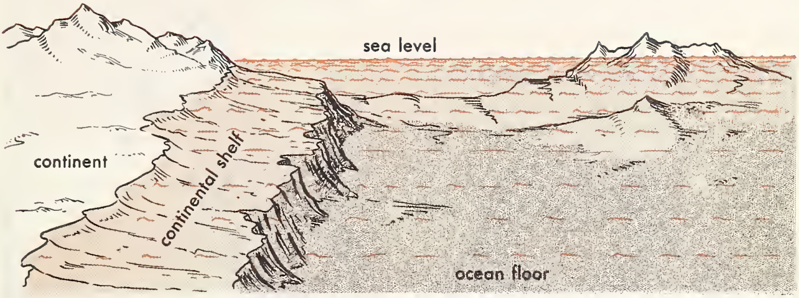


Fig. 2-2. A continental shelf is a submerged, gently sloping plain around the shores of every continent.

taking soundings was laborious and time-consuming; sometimes it took many hours to make a single reading. In 1920 all this was changed by the invention of sonar, a "searchlight" that uses sound waves. Sea water transmits sound much better than air, and a sonic sounding instrument measures the time it takes a sound pulse to travel to the ocean floor and back again. From this time interval the depth of the ocean can easily be computed. Indeed, the instrument actually plots a profile of the sea floor as the ship moves along. The result is that a vast number of soundings have now been made in all the oceans. What have these soundings revealed? Strange as it may seem, the ocean basin consists of three entirely different and separate regions: (1) the continental shelf, (2) the continental slope, (3) the floor itself.

The Continental Shelf. Around the shores of every continent there is a submerged margin called the continental shelf. This shelf is a gently sloping plain extending seaward ten miles or so off some coasts, but as much as 400 miles off others. If the coast is mountainous, the shelf is relatively narrow; if the coast is low lying the shelf is much wider. For example, off the Pacific Coast, near the site of recent mountain building, the shelf is only 20 miles wide. But off the east coast the shelf extends much farther and at the Grand Banks of Newfoundland it is rather more than 250 miles wide.

The waters over the continental shelf are shallow; they vary in depth from about 200 to 600 feet. The plain itself is covered with material such as sand and soil, derived from the land and carried there by running water. The world's great fisheries are confined to these shallow waters, and, off the coasts of California, Texas and Louisiana, the shelf is yielding large amounts of oil.

Actually, the continental shelf is a transition area between land and ocean; it is really a continuation of the land mass. In past ages many parts of the shelf have been alternately flooded and dry depending on the rise or fall of the sea. Indeed, the edge of the shelf is supposed to mark the lowest level reached by the seas during that Ice Age when vast quantities of water were piled up as ice on the land.

The Continental Slopes. A sea basin really begins at the continental slopes. The slopes are steep (the sea floor drops about 400 feet per mile) and extend for 50 miles or so beyond the edge of the shelf, and plunge to great depths, often more than 15,000 feet.

The Ocean Floor. The term floor is misleading since it implies a relatively flat area whereas, as already stated, the ocean floor is far from flat. The physical

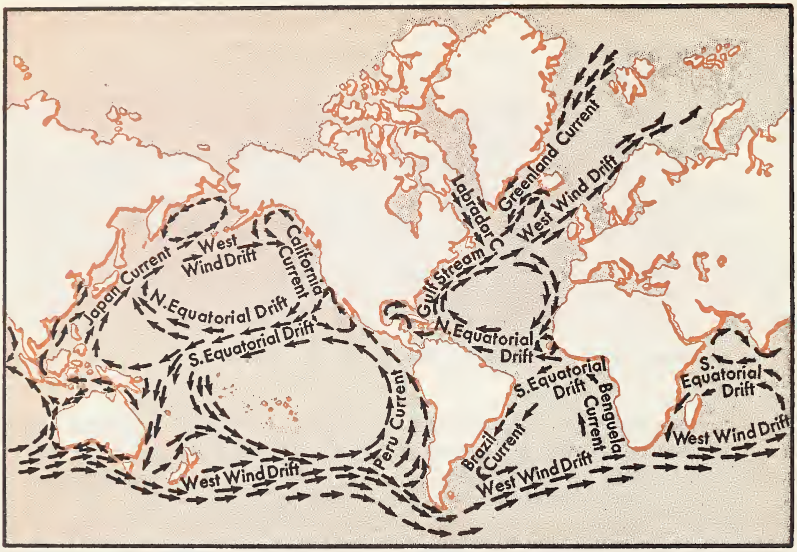


Fig. 2-3. Steady currents circulate through the waters of the oceans. Notice the Gulf Stream which extends from the Gulf of Mexico to the northwestern shores of Europe.

features, as on land, include plains, plateaus, mountains, valleys and even canyons.

Submarine mountains are probably the most interesting feature, and the range that has been studied most thoroughly is the Atlantic Ridge. This is the most stupendous mountain range on earth, far higher and wider than the Rockies. Running north-south in mid-Atlantic, it is almost 10,000 miles long. It extends southward from Iceland, crosses the Equator, and continues as far south as latitude 50° S. If you follow these directions on the map on page 38 you will get an idea of the immensity of the Atlantic Ridge. Its highest peak rises 27,000 feet from the sea floor and emerges 8000 feet above sea level. This peak is Pico Island in the Azores.

The floor of the Pacific, though not as mountainous as that of the Atlantic, is studded with immense volcanic cones. It too has a number of mountain chains, one being a submerged volcanic ridge that stretches almost 2000 miles between the emerging peaks that form the Hawaiian Islands and Midway Atoll.

Marine mountains, unlike land mountains, are mostly of volcanic origin. This is particularly true of the Hawaiian Islands, where violent volcanic activity continues to this day. Mauna Kea in this group is the highest mountain on earth. It rises more than 30,000 feet above the sea floor and nearly 14,000 feet above the surface of the Pacific. It is impossible to imagine the gigantic forces involved when a volcanic mountain is formed in deep water. In land volcanoes, the lava is simply ejected into the air. But the lava that formed the submerged Pacific ranges was forced up in spite of the enormous pressure exerted against it by thousands of feet of water.

Before leaving the sea floor reference should again be made to the deep valleys and gorges which occur. These are almost always found near the continental slopes and not within the basin floor itself. Into these chasms ocean currents have deposited sediments carried from the land so that sometimes the accumulations are more than a mile deep. In Chapter 8 we shall refer again to these sediments and discuss their role in mountain building.

Ocean Currents. The waters of the oceans are not stagnant, but are traversed by steady currents. These currents may flow from north to south, from east to west, or from the surface to the depths. They affect the climate of continents; they mix the waters of oceans so that no water can be thought of as solely Atlantic or Pacific.

What causes ocean currents? The forces which drive water in steady streams within water are complex but at least two factors should be mentioned. These are (1) changes in the density of sea water, (2) prevailing winds.

Currents Due to Density Changes. The surface of the sea is not heated equally by the sun; equatorial waters are warmer than waters in higher latitudes. What is the effect of this unequal heating? As water is warmed it expands and becomes lighter, or less dense; as water is cooled it contracts and becomes heavier, or more dense. Cold water, therefore, sinks below warmer water and a current may be started in this way. Indeed, a current of this sort causes a slow exchange between polar waters which creep along the ocean floor and equatorial waters which flow along the surface.

Then again "vertical" currents are also caused by differences in salinity. In the warm regions of the Indian Ocean, for example, a great deal of evaporation takes place from the surface waters, leaving behind a denser salt solution that tends to sink. On the other hand, there is much less evaporation in the cold water near the antarctic so that its water is not as dense as that of the Indian Ocean. As a result, denser water from the Indian Ocean slides down at a steep angle below the colder antarctic waters. Many vertical currents are formed this way in the oceans.

Currents Due to Prevailing Winds. Prevailing winds, which are considered in Chapter 11, are the main cause of surface currents. The steadiest winds are the Trade Winds which blow diagonally towards the Equator from the north-east and south-east. These winds drive equatorial currents with them in a westerly direction as shown in Fig. 2-3. The rotation of the earth from west to east deflects the current so that it curves clockwise in the Northern Hemisphere and counterclockwise in the Southern Hemisphere to form easterly currents. These currents are the Gulf Stream in the North Atlantic and the Brazil Current in the South Atlantic.

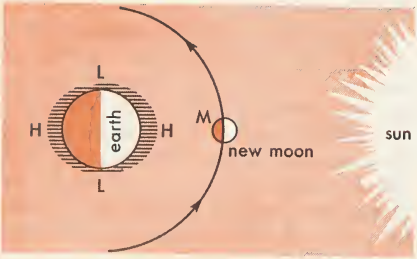
The Gulf Stream, the best known of all ocean currents, is derived from a warm equatorial current. It turns in the Gulf of Mexico (from which it derives its name) and then flows north along the east coast as far as Cape Hatteras. Then, impelled by the earth's rotation it flows northeast across the Atlantic, eventually reaching the western shores of Europe. The Gulf Stream is about 50 miles wide, about 1500 feet deep, and it travels about five miles per hour. This warm broad stream transports through the Atlantic about 70 million tons of water per second, an amount which exceeds by a thousand times, the water discharged by the Mississippi into the Gulf of Mexico. No wonder the Gulf Stream warms the shores of France, England, and even Norway. Have you ever noticed that the British Isles lie in roughly the same latitude as much of Labrador? This is difficult to realize when we compare the mild British winters with the bitterly cold winters of northern Canada. This difference in climate is attributed largely to the Gulf Stream.

Tides. The periodic movement of the tides is the most impressive motion of the sea. The tides are something we can see and measure; twice every day they rise and fall; and they set every part of the ocean in motion. Near the land, the rise and fall of the sea can easily be measured by a tide gauge. Out to sea, however, the tides cannot be seen and cannot be measured in this way. Instead they are measured by a pressure gauge, a kind of barometer (page 25), placed on the sea floor. What causes tides? The answer to this question involves the force of gravitation.

Gravitation and Tides. The ***law of gravitation*** says that ***all bodies in the universe attract each other.*** The heavier the bodies, the greater is the force of attraction between them; the greater the distance between the bodies, the less is the attractive force between them. Thus the moon attracts the water of the seas and, consequently, water piles up under the moon. In this way, a ***high tide*** is formed every day when the moon is overhead. The moon also pulls on the earth. As a result another high tide is formed on the far side of the earth, opposite to that on the near side.

As the earth rotates on its axis, different places on its surface are brought under the moon so that they, in turn, experience high tide. If the moon did not move, every place on the coast would have two high tides with every rotation of the earth; that is, two high tides every 24-hour day. But the moon revolves around the earth in the same direction as the earth spins on its axis. How does this affect the time between high tides? Because of the motion of the moon, a place on the earth must go more than a full circle to come under the moon again. This extra rotation takes 50 minutes. Hence there are two tides every 24 hours, 50 minutes; or, stated another way, high tide occurs at intervals of 12 hours and 25 minutes.

Fig. 2-4. Spring tides are the highest tides. They occur when earth, moon, and sun are in line.



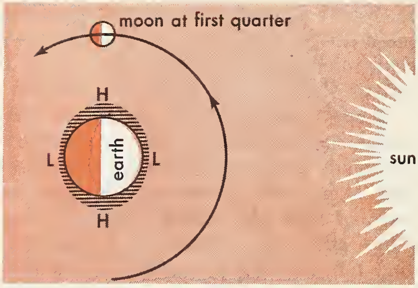


Fig. 2-5. Neap tides are lower than normal tides. They occur because the pull of the sun tends to neutralize the pull of the moon.

The sun also causes tides but, because it is so far away, it is less effective than the moon. Indeed, the gravitational pull of the sun on the ocean is only half as great as the moon's pull. Tides are therefore determined by the relative positions of moon ***and*** sun. When sun and moon pull at the same time in the same direction, tides are highest and are called ***spring*** tides. Spring tides occur at new or full moon because in these positions, earth, moon and sun are in line. Spring tides are the times, twice every month, when the water rises highest in the harbors. At the first and third quarters of the moon, earth, moon and sun are not in line; they are at the corners of a triangle. In these positions, the pull of the sun tends to neutralize the pull of the moon, and the tides are lower than normal. They are called ***neap*** tides. How many neap tides are there every month?

Other Factors Affecting Tides. The tidal picture throughout the world is not as simple as the one described above. There are factors which cause complications. We know, for instance, that places on both shores of the Atlantic usually have two tides a day; in some parts of the Pacific there is only one tide a day, and this is also true of some places on the Gulf of Mexico. Then again, the vertical drop between high and low tide varies at different places. At Nantucket Island off the Massachusetts shore the variation is only a foot or two; in the Bay of Fundy (Nova Scotia) it is 40 to 50 feet. Yet both these places are in the same body of water, namely the Gulf of Maine. Still another example concerns the Panama Canal. At the Pacific end of the Canal the tide lifts 15 feet or so; at the Caribbean end, only 40 miles away, it lifts only a foot or two. How can we explain these differences?

It is evident that there are some controlling factors besides the sun and moon, and no geographer would claim he has the complete answer to this complex tidal problem. All are agreed, however, that the sun and moon set the water in motion, but ***how far*** it will rise depends upon the shape of the basin and the slope of the bottom.

The ocean floor consists of many natural basins and the tidal rhythm is caused by the rocking of the water up and down in each particular basin. Suppose such a basin is represented by water in a bath tub which we rock up and down about the middle. The water rises and falls in "tides" and the height of the tide depends on the length of the tub, the amount of water in it, and the rhythm of the swing. The highest tides are at the ends of the tub and in the middle there is little or no tide.

The rhythmic swing due to the moon and sun rocks the water back and forth every 12 hours or so. If the swing due to the moon reinforces the swing due to the basin, extremely high tides are formed. Thus it is thought that the Bay of Fundy lies at the end of a natural oscillating basin with a time of swing of about 12 hours. This coincides with the period of the ocean tide caused by the moon and sun. Still another factor favoring the formation of unusually high tides in the Bay of Fundy is that the water is funnelled into a converging estuary. It is also thought that Nantucket Island is somewhere near the middle of the basin, which would account for its unusually small tides.

Things to Remember

The human body is composed of about 70 per cent water.

Sea water contains about 3.5 per cent of salt.

A ***solution*** is a mixture of water and a dissolved substance.

A ***saturated solution*** contains as much of any one solid as it can dissolve.

As the temperature is raised the greater is the ***solubility of a solid*** in water.

Prevailing winds in Canada are westerly.

The ***largest oceans*** are the Atlantic and Pacific.

The ***continental shelf*** is the gently sloping submerged margin of every continent.

Ocean currents are formed chiefly by prevailing winds.

Tides are caused by the gravitational attraction of the moon and sun on sea water.

Questions

GROUP A

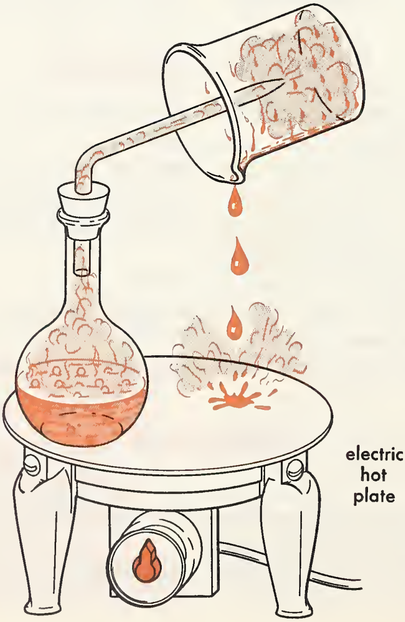
1. The first rains of the earth probably fell continuously for many years. Why?
2. What is a solution? Give an example.
3. The Great Salt Lake in Utah contains almost 30 per cent salt. What does this mean?

4. A solution weighing 50 grams is evaporated to dryness and the salt residue weighs 4.5 grams. Find the percentage of salt in the solution.
5. "The dense forests of British Columbia are due to rains from the Pacific." Explain what this means.
6. The melting of polar ice would cause sea level to rise. Why?
7. What is a sonic sounding instrument?
8. What is the similarity between Pico Island in the Azores and Mauna Kea in the Hawaiian Islands?
9. What is the direction of the trade winds?

GROUP B

10. What are the common ideas of scientists concerning the formation of the first oceans?
11. A solution is made by dissolving 4.5 ounces of salt in 50 ounces of water. What is the percentage of salt in the solution?
12. What is meant by the Continental Shelf? Compare the continental shelves off the Atlantic and Pacific coasts of Canada.
13. Write briefly on the physical features of the ocean floor.
14. What is the Gulf Stream? How is it formed, and in which direction does it flow?

Fig. 2-6. When rain fell upon the surface of the hot earth, it instantly vaporized.



15. Explain how (*1*) high tides are formed every day and (*2*) spring tides are formed.
16. The greatest tides on earth are in the Bay of Fundy. How can you account for this?

Things to Do

Cloud formation and rain · Half fill a Florence flask with water. Close the flask with a stopper and bent glass tube, the end of which has been constricted as in a medicine dropper. (See Fig. 2-6.) Place the flask on an electric hot plate. In a few minutes the water boils and steam rushes out of the tube.

Hold a liter beaker in the path of the steam. The steam condenses to droplets (forming a cloud). Hold the beaker in a wet, cold towel. The droplets are further cooled and they coalesce to form drops (rain) which fall from the spout of the beaker.

Direct these drops onto the hot plate and notice that they vaporize instantly. The falling of drops of water on the hot plate simulates the conditions when the first rains fell upon the hot surface of the earth.

3

THE EARTH'S ATMOSPHERE

ALTHOUGH we cannot see air, there is plenty of evidence to prove it is a real substance, just as real as a solid. For example, air in motion can drive a sailing ship across the ocean, it can turn a windmill, it can buoy up a bird or plane in flight. Unless air were a substance it could do none of these things. This substance, air, is essential to all forms of life; neither plants nor animals could exist without it. Nor could fuels burn without air. Automobiles, locomotives, and even jets would lose their power of locomotion if there were no air to burn the fuels in their engines.

What Is Air? Air is a mixture of five gases: nitrogen, oxygen, argon, carbon dioxide and water vapor. Its approximate percentage composition is nitrogen 78%, oxygen 21%, argon 1%, carbon dioxide .04%. Notice that about four-fifths of the air is nitrogen, whereas there are only 4 parts of carbon dioxide in 10,000 parts of air. Notice also that water vapor is left out of the percentage list. This was done because the percentages of the gases in the list remain fairly constant, but the percentage of water vapor varies considerably. On a very dry day there may be only a trace of water vapor in the air, but on a damp day there may be as much as 4 per cent of it.

Of these substances nitrogen, oxygen and argon are ***elements;*** carbon dioxide and water vapor are **compounds.** What are elements and what are compounds? These are important questions.

Elements and Compounds. ***An element is a substance which contains only one kind of matter;*** it cannot be decomposed into simpler substances by chemical means. The number of elements in nature is relatively few; there are only 90 of them. A few, 12 in all, are man-made; they do not occur in nature. Hence the total number of known elements is 102. Elements, at ordinary temperatures, may be solids, such as carbon and silicon; or liquids, such as mercury and bromine; or gases, such as oxygen and hydrogen.

By far the most plentiful elements in the earth as we know it (that is, in the air, the sea, and the earth's crust) are the gas, oxygen, and the solid, silicon. Oxygen accounts for more than 50 per cent of the weight of the known "earth" and silicon for more than 25 per cent of it. We must not think, however, that oxygen and silicon always occur in nature as elements, or as we usually say, in the free state. They are usually combined with other elements as compounds.

Compounds are formed whenever two or more elements combine chemically.

There are far more compounds than elements, hundreds of thousands of them. The most abundant compound, and the most familiar, is water, a compound of hydrogen and oxygen. Another important compound is the gas, carbon dioxide, a compound of carbon and oxygen. The most abundant compound in the earth's crust is silicon dioxide, or silica, as it is usually called. This compound of silicon and oxygen makes up about 60 per cent of the earth's crust; it is a constituent of most of the rocks of the crust.

Elements and compounds can be represented by symbols and formulas respectively. These are a convenient form of shorthand for the chemist. The symbol for the element oxygen is O, for hydrogen H, for the element nitrogen N, and for carbon C. The formula for water is H_2O; the formula for carbon dioxide is CO_2, and for silica, SiO_2. The formula H_2O means that one molecule of water (or water vapor) is made up of two atoms of hydrogen and one atom of oxygen. Here we have used two new terms, ***atom*** and ***molecule.*** What do they mean? We shall use these words frequently in later chapters, and it may be well to discuss them at this time.

Atoms and Molecules. We must not think of oxygen in a bottle as completely ***filling*** its container. The oxygen is made up of minute particles which move at random at high speeds. The same is true of a bottle containing nothing but water vapor; its particles also move about rapidly in all directions. There must be a good deal of empty space in the bottle, otherwise the particles would have nowhere to go. These particles are called ***molecules.*** What then is a molecule? One answer to this question would be, ***A molecule is the smallest particle of an element or compound that can exist by itself.*** Notice the words ***by itself*** at the end of the sentence. These are important because it would be quite wrong to say, A molecule is the smallest particle of an element that can exist. Let us see why this is so.

The molecules of oxygen are composed of two parts which move together as a pair. These parts were originally separate but, for reasons we shall discuss later, they combined to form a molecule. Molecules are therefore not ***ultimate*** particles; they have been formed from still smaller particles by chemical action. These smaller particles are called ***atoms.*** Can you now say what an atom is? One answer to this question would be, ***An atom is the smallest particle of an element that can take part in a chemical reaction.*** We can therefore represent an atom of oxygen by the symbol O, or, simply by O; and an atom of hydrogen can be represented by the symbol H. That is, the same symbol is used for an element and for a single atom of the same element. The formula for an oxygen molecule is O_2, for a hydrogen molecule H_2 where the subscript 2 indicates the number of atoms per molecule. The formula for a molecule of water vapor is H_2O and the formulas for the molecules of carbon dioxide and silica are CO_2 and SiO_2 respectively. (Notice that the symbol for the element Silicon is Si. This is because the symbol S is used to represent an atom of sulfur.) How many atoms are there in a molecule of carbon dioxide? How many atoms of silicon are there in a molecule of silica?

The molecules we have considered so far consist of not more than three atoms. We must not think, however, that all molecules are as simple as these. Some of them, as we shall see later, may contain 10, 20, and even more atoms.

Gas Molecules are Always Moving. Earlier in the chapter, we read that gas molecules move at high speeds. Let us devise a simple experiment to prove this. Since we wish to ***see*** the gas, our choice is limited and we shall use liquid bromine for this purpose. But we must use it with care because liquid bromine on the hands will cause a severe burn. Even bromine vapor is poisonous; it should not be inhaled. Fortunately it is red in color and can easily be seen.

Demonstration

3-1. To show that gas molecules move

Place a bottle of bromine on a piece of cardboard about a foot square. Remove the stopper and cover the bottle with a large inverted battery jar. Leave the bottle in this position for half an hour or so.

What do you observe? A red vapor rises from the liquid bromine and, in 10 or 15 minutes, the jar is completely filled with it.

What conclusions can be drawn from our observation? First of all, liquids evaporate to form a vapor or gas. Secondly, ***gas molecules move in all directions,*** otherwise the color could not spread throughout the jar. There are two other statements we should make concerning gas molecules, although we have not proved them from the above experiment. These are (1) ***the higher the temperature, the more rapidly do gas molecules move*** and (2) ***at the same temperature lighter molecules move faster than heavier ones.*** You could probably think of a simple experiment to prove the first point. Concerning the second, let it be said that molecules of hydrogen move faster than molecules of bromine vapor because the hydrogen molecules are lighter.

With this background of elements and compounds, atoms and molecules we are ready to ask, and to try to answer, the questions, Where does the atmosphere come from? Why does it not escape into space?

Air Is a Mixture. Earlier in the chapter we stated that air is a mixture of five gases, and that almost 99 per cent of dry air is oxygen and nitrogen. Since these gases are not chemically combined, it ought not to be a difficult matter to separate them. This separation is made easier because oxygen is chemically active and takes part in many reactions, whereas nitrogen is chemically inactive and does not readily react chemically. An example of this difference is demonstrated in the rusting of iron. Iron rusts in damp air to form a compound called iron oxide; nitrogen does not. We shall make frequent references to chemical activity in subsequent chapters.

The chemical action between iron (or steel) and oxygen can raise some vexing problems. Automobiles, bridges, and even skyscrapers are constructed of steel. If the metal were exposed to air it would rust, and disintegration would set in. To prevent this, the exposed surfaces are painted, or enamelled or, in some cases, plated with another metal (such as chromium) that does not readily rust.

How the Earth Retains its Atmosphere. We live at the bottom of an ocean of air several hundred miles in depth. Although the earth travels round the sun at the enormous speed of nearly 70,000 miles an hour, the air stays on the earth. Why is this? At first sight it might seem that the earth would leave the air behind, somewhere in space. The atmosphere is retained because the earth attracts the air molecules with a force great enough to prevent them from escaping. If the earth were smaller in size, the force of gravitation would be less and the air molecules would be more difficult to hold in place. There is no atmosphere on the moon. But the moon is much smaller than the earth, and its gravitational pull is not large enough to keep gas molecules near its surface. For this reason the moon's atmosphere, if there ever was one, escaped into space.

Origin of the Earth's Atmosphere. The earth's atmosphere is quite different from the atmospheres known elsewhere in the solar system. Oxygen, carbon dioxide, nitrogen, and water vapor are much more abundant on the earth than on other planets. Oxygen, in particular, is almost completely absent on them, even on Mars. Where does the atmosphere come from? When scientists try to answer this question they are in the realm of speculation, but there are sound reasons for their guesses.

The most plentiful elements in the universe are hydrogen and helium, the two lightest elements. These two elements account for 99 per cent of the substance of the universe, hydrogen being ten times more abundant than helium. It is therefore reasonable to suppose that the earth's original atmosphere contained these gases in large amounts. However, the earth's gravitational pull on these light elements is so slight that, like the moon's atmosphere, they escaped into space.

The earth probably got its second atmosphere as the molten rocks were cooling and crystallizing. At that time, gases were released as rock compounds were being formed in the earth's interior. This view is supported by examining the gases released in volcanic eruptions that occur today. In one of the active Hawaiian volcanoes the composition of the released gases is as follows: water vapor 68%, carbon dioxide 13%, nitrogen 8%, and the rest is sulfurous fumes. This then may have been roughly the composition of the atmosphere when the earth's surface was still hot. As the earth cooled and the water vapor condensed to form the seas, the percentage of water vapor would be drastically reduced and the relative percentages of nitrogen and carbon dioxide thereby increased.

But so far no oxygen. Where does it come from? Oxygen did not enter the atmosphere until plants appeared. By a process called photosynthesis all green plants live by using up carbon dioxide and forming oxygen. As plant life evolved through the ages, the percentage of carbon dioxide was reduced and the oxygen in the atmosphere slowly increased so that, eventually, there was enough oxygen in the air to support animal life.

Animals, of course, inhale oxygen and exhale carbon dioxide, the converse of the photosynthesis operation. And so a balance was reached between these two opposing operations which support plant and animal life. When this balance was reached the earth had got its third or present atmosphere. And scientists think that our present atmosphere is about the same in composition as it was five hundred million years ago.

Air Exerts Pressure. Although air is a gas, the total weight of the earth's atmosphere amounts to the stupendous figure of almost five thousand billion tons (5,000,000,000,000 tons). This weight is difficult to imagine, but it would be about the same weight as a rectangular block of granite, 1000 miles long, 1000 miles wide, and one mile high.

Air is easily compressed, as we realize, when we use a bicycle pump. Therefore, the densest air is at the earth's surface and it gets progressively thinner as altitude increases. At an altitude of about 3½ miles, the pressure of the air is just about half the pressure at the surface. This is because the molecules are twice as far apart 3½ miles up as they are at sea level. If you take a deep breath at this altitude, only half as much air goes into your lungs as at sea level, and, to take in the necessary amount of oxygen, you have to breathe twice as fast. Six miles up air is too thin to support life; the mountain climbers who conquered Everest in 1953 were all equipped with an oxygen apparatus. Actually, the atmosphere probably extends 250 miles or so above the earth's surface.

It is not surprising therefore that this heavy atmosphere exerts a considerable pressure upon all bodies on the earth's surface. This can be shown by taking the air out of a tin can.

Demonstration

3-2. To show that air exerts pressure

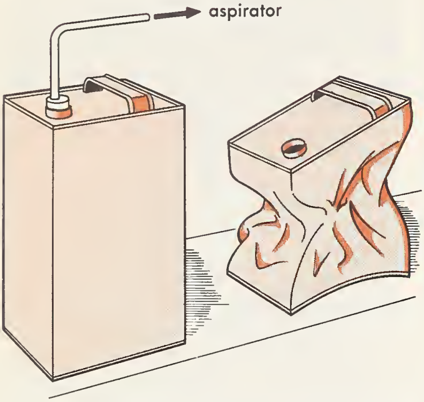
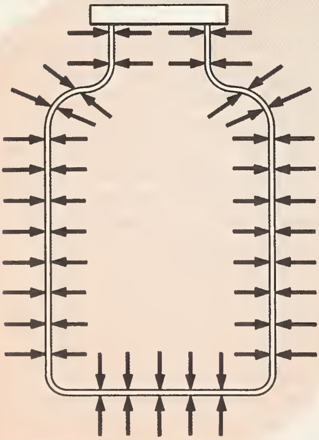


Fig. 3-1. If air is taken out of a tin can, it buckles and collapses.

Place a short piece of glass tubing through a one-hole stopper, and fit the stopper into the outlet of a one-gallon tin can. Attach the glass tube to an aspirator and turn on the faucet. In a few minutes the can begins to buckle, and eventually the can collapses completely, as shown in Fig. 3-1.

Fig. 3-2. In a closed container, under ordinary conditions, the *outward* pressure of the inside air balances the *inward* pressure of the outside air.



What caused the can to collapse? Clearly, the pressure of air exerted *inwards*. But why didn't the can collapse before the air was withdrawn? The only explanation can be that air exerts a pressure in all direction. As a result, the *outward* pressure exerted by the air inside the can normally balances the inward pressure of the outside air, as shown in Fig. 3-2. If the air is withdrawn from the can, the pressure of the air outside acting against the can exceeds the pressure of air remaining inside. The can is not strong enough to resist this inward thrust.

Atmospheric pressure is, of course, exerted on our bodies just as it is on the tin can. But we are not crushed by it, nor are we even aware of it, because we inhale air. The counter pressure of the inhaled air is transmitted by the blood and tissues of the body and resists the inward pressure of the atmosphere. Let us consider another example. Suppose the pilot of a pressurized jet plane, travelling through the stratosphere, were suddenly ejected from his plane. What would happen to him? The pressure inside his body would be much greater than the outside pressure of the thin air, and he would probably burst like an inflated balloon.

How can atmospheric pressure be measured? This is usually done by an instrument called a *barometer*, a word meaning a *measure of weight.*

Demonstration

3-4. To measure pressure with a barometer

Take a barometer tube (which is a glass tube, closed at one end, about a yard long and about half an inch in diameter) and completely fill it with mercury. Close the open end with the finger, invert the tube, and place its open end well below the surface of the mercury in a small basin or beaker. Now remove the finger.

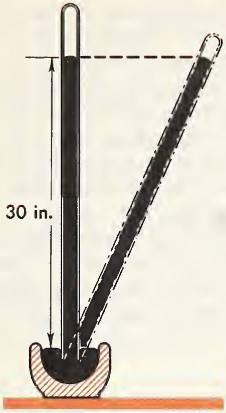


Fig. 3-3. A simple barometer.

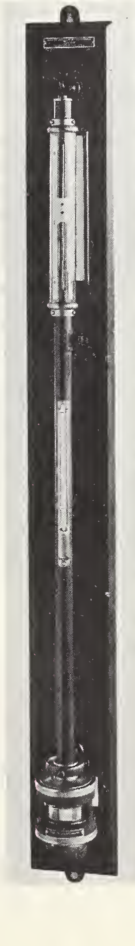
Notice that, as soon as the finger is removed, some mercury falls out of the tube. Tilt the tube as shown in Fig. 3-3 and notice that the vertical height of the mercury level in the tube remains steady. Now restore the tube to the vertical position and hold it in place with a clamp and stand. Measure the height of the column, in inches and also in centimeters, from the surface of the mercury in the basin to the top of the column.

How does this mercury barometer operate? It depends on a balance between the pressure exerted by the mercury and the pressure of the air. The downward pressure of the mercury in the tube tries to force the mercury out of the tube; the downward pressure of the atmosphere exerted on the mercury surface in the basin tries to force mercury *up* the tube. When these two forces balance each other, the mercury remains at a constant level. This level, if we ourselves are near sea level is probably about 30 inches (30 in.) or 76 centimeters (76 cm.) above the outside surface. In other words, we usually express atmospheric pressure in terms of the height of the mercury column supported by the atmosphere, and we can say that ***the pressure of the atmosphere equals the pressure of a vertical column of mercury about 30 inches or 76 centimeters long.***

When the tube was tilted, the mercury level remained at the same vertical height but the space above the mercury became smaller. Indeed, if the tube is sloped sufficiently the space can be made to disappear altogether. This proves that the space does not contain air; it is empty. An empty space is called a vacuum.

Why is mercury, rather than water, used in a barometer? Mercury is a heavy liquid; it is 13.6 times as dense as water. Hence if water were used its level would be 13.6 times higher than the mercury level; that is, the height of the water column would be about 34 feet. (How is this figure obtained?) Hence, mercury is a more convenient barometric liquid than water.

Fig. 3-4. The mercury level is read by a scale permanently attached to an accurate barometer.



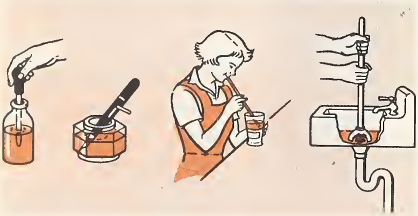
The pressure of the air does not, of course, remain constant day after day. Variations in air pressure cause the mercury column to rise or fall through small distances. Barometric readings are closely associated with weather; they are a part of the data needed in making weather predictions, a topic discussed in Chapter 13.

There are other ways of measuring air pressure besides referring to the height of a mercury column. Sometimes we measure it in terms of the force exerted upon one square inch of surface. For example, a pressure equivalent to 76 cm. of mercury is also equal to about 15 pounds per square inch. Expressed in these terms, we can say that ***the pressure of the atmosphere is approximately 15 pounds per square inch.***

Units of Length. You will recall that we measured the mercury column in inches and in centimeters. These are units in the British and metric systems of length respectively. As you know, both systems are used and it may be well to know how to convert a length in one system to the corresponding length in the other. Suppose the barometer reads 29.8 in. and 75.7 cm. Therefore, in length, 29.8 in. equals 75.7 cm. Hence, 1 inch equals $\frac{75.7}{29.8} = 2.54$ cm. It is helpful to remember that 2.54 centimeters equal one inch.

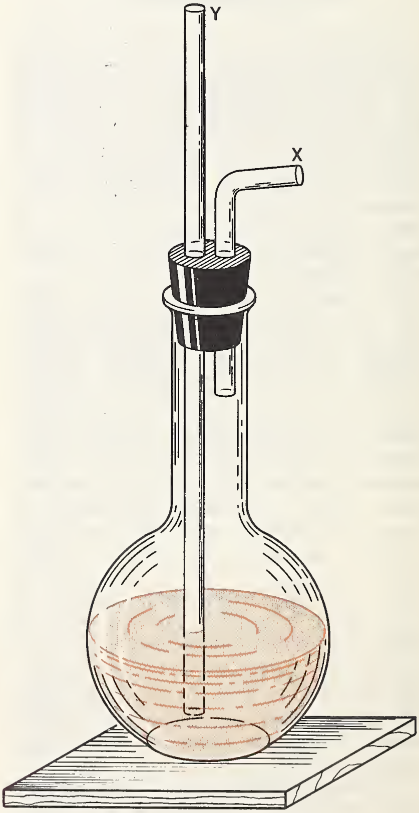
Applications of Air Pressure. A vast number of devices depend upon air pressure for their operation. Among them are the aneroid barometer, water pumps and the carburetor in an automobile engine. At this time, however, it may be well to consider a few of the more familiar devices. Why, for instance, does the liquid enter your mouth when you suck on the straw in the milk shake? By sucking you take air out of the straw and so diminish the pressure within it. As a result, the air pressure exerted on the liquid surface forces the milk shake into your mouth. Now suppose you place an air-tight cap over the tumbler, admit the straw through a hole in the cap, and seal the cap to the straw with paraffin wax. Again suck the air out of the straw. What happens this time? Why?

Fig. 3-5. Some applications of air pressure.



You now ought to be able to explain how a medicine dropper works, and how a fountain pen is filled with ink. When the small lever on the fountain pen is turned outward, the other end of the lever inside the pen compresses a rubber bulb, which in turn expels air bubbles through the ink. When the lever is turned back toward the pen, air pressure forces ink into the bulb. Then again, when you wish to pour liquid out of a sealed tin can it is best to make two openings on opposite sides of the top of the can. Why is this?

Fig. 3-6. A device to show gas pressure.



Or, again, a hemispherical rubber plunger on the end of a stick is often used to clear a blocked drain pipe. How does this work? The plunger, placed over the end of the drain pipe, is deflated several times by pushing down on the stick. Air in the drain pipe is thereby forced out at the base of the plunger, and atmospheric pressure, transmitted through the water in the drain, is able to remove the obstacle.

Finally, let us make the simple device shown in Fig. 3-6 and see how air pressure affects its operation.

Demonstration

3-4. A device to show gas pressure

(*a*) Place the end of the tube X in the mouth and suck air out of the flask. Bubbles of air enter the water through tube Y. Why is this?

(*b*) Now, with the mouth still at X, blow air into the flask, and quickly close the end X with the finger. Notice that water has risen in the tube Y. How do you explain this?

(*c*) Place a short rubber tube over the end X and seal it with a pinch clamp. Now blow down the end of tube Y and then take the tube out of your mouth. What happens? Why?

(*d*) Release the pinch clamp. What happens? Why?

Air Currents. The atmosphere is not stagnant over the earth, but, like the turbulent oceans, currents flow through it. Air is more compressible than water so that air currents respond more readily than water to changes of temperature. Air currents, or winds, are caused primarily by the sun. For example, near the equator, the hot air rises. In rising, it cools, and in cooling, it loses much of the water vapor which condenses as rain. It is the heavy annual rainfall in equatorial regions that accounts for the lush vegetation usually found there.

Winds are caused by a flow of air from a region of high pressures to a region of low pressure. How can we account for these areas of high and low pressure? If a part of the earth's surface is strongly heated, the air above it is warmed and therefore expands and becomes less dense. This produces low pressure. On the other hand, air over a cold surface becomes more dense and produces high pressure.

An area of high pressure, or a "high" as it is called, will form over the cool ocean, while a "low" will form over a strongly heated adjacent continent. There are large highs and lows that stay pretty much in one place, but there are also smaller highs and lows that wander over the earth's surface in an irregular manner. As a result, the winds of the earth are complex and ever changing.

Things to Remember

The approximate **composition of air** is nitrogen 78%, oxygen 21%, argon 1%.

An **element** is a substance that contains only one kind of matter.

Compounds are formed whenever two or more elements combine chemically.

A **molecule** is the smallest particle of an element or compound that can exist by itself.

An **atom** is the smallest particle of an element that can take part in a chemical reaction.

Gas molecules move in all directions and the higher the temperature the faster they move.

The **most plentiful element** (*a*) in the universe is hydrogen, (*b*) on the earth is oxygen.

Normal **atmospheric pressure** at sea level is equivalent to the pressure of a mercury column 76 centimeters (or about 30 inches) in height.

Atmospheric pressure also equals 15 pounds per square inch.

One inch equals 2.54 centimeters.

Questions

GROUP A

1. State the approximate composition of air.
2. Define the terms **element** and **compound** and give three examples of each.
3. Name the most abundant, (*a*) element and (*b*) compound in the earth.
4. There is no atmosphere on the moon. Why not?
5. At an altitude of about 3½ miles, atmospheric pressure is approximately halved. Why?
6. Compute the height of a water barometer in feet if a mercury barometer reads 29 inches.
7. Compute the barometric reading in inches if the height of the mercury column is 77.5 cm.
8. Compute the barometric reading in centimeters if the height of the mercury column is 29.0 inches.

GROUP B

9. (*a*) Define the terms **atoms** and **molecule.** (*b*) Illustrate the difference by reference to the element oxygen.
10. If the stopper is removed from an "ammonia bottle," the smell of ammonia soon spreads to all parts of the room. Why is this? Explain.
11. Describe an experiment to prove that oxygen is an active ingredient of the air.
12. Explain why the earth lost its atmosphere of hydrogen but gained an atmosphere containing oxygen.
13. Draw a mercury barometer and explain how it operates.
14. Explain clearly how a medicine dropper works.

Things to Do

To record barometric readings and weather conditions · Make a barometer as explained in Demonstration 3-3. Read the barometer at approximately the same time every day for two or three weeks. Record on a chart (*a*) the date, (*b*) the barometric reading in centimeters, (*c*) the kind of weather, such as fair, rain, windy, etc.

After 3 weeks or so examine the data you have collected. Is there any relationship between the barometric readings and the weather? For example, is rain or wind associated with a rising, falling, or steady pressure?

Earth, Abode of Man

LOOKING AHEAD

If we travel by plane from Newfoundland to Vancouver Island, we see many physical features of the country (see Fig. 8-15). As we climb we can see the Grand Banks, the shallow continental shelf of the ocean, which extend nearly 300 miles off the east shore of Newfoundland. We pass over the very rugged terrain of Newfoundland: it is part of the Appalachian belt which extends across New Brunswick and the Gaspe Peninsula into the eastern United States. This belt consists of low mountains with weathered peaks of sandstone and limestone that do not look very formidable. As we head west we pass over the mouth of the St. Lawrence, which has cut a deep submarine canyon in the continental shelf. Over the mainland now we see to the north the broad extent of the Canadian Shield. This is a very large geological division, almost one half of Canada. The land here is very rough and tree-covered, with small lakes.

We continue along the St. Lawrence lowlands, across the Great Lakes, and soon, just east of Winnipeg, we see the Canadian Shield give way to the broad flat lands of the Great Plains. The plains rise gently as we travel west, from an elevation of 700 feet at Lake Winnipeg to over 3400 feet at Calgary. Now we see the huge cliffs of the Rockies, with peaks over 13,000 feet high. West of the Rockies and the Columbia Mountains we are flying over a plateau which has been cut by deep canyons. When we reach the Pacific Coast we meet the granite peaks of the Coastal Range and see the picturesque long bays or fiords which were cut by the ice sheets as they advanced to the ocean. Finally, we pass over the Straits of Georgia to Vancouver Island, the largest island rising from the Pacific Continental Shelf.

Geologically, this is an interesting trip. Why do the mountains run from north to south? Why do we find rugged peaks in the west and rounded hills in the east? Why do we find flatlands in the prairies and igneous rocks in the Coast Range?

Every exposed rock tells its own story, and together the stories reveal the biography of the earth. The earth as we see it today is only one picture of an ever-changing pattern. To be sure, the change is slow, but nevertheless it is continuous. There is no doubt that much of the stuff mountain peaks are made of once formed the floor of an ocean. Indeed, fossils, which are a record of life on the earth, have been found on mountain tops, indicating that the rock of the mountains was formed from material deposited in an ocean.

← *Young mountains are high and rugged, not rounded by erosion. Their peaks are jagged, their slopes steep and treeless.* (Ewing Galloway)

4

THE EARTH—ITS SHAPE AND MOTIONS

When early man looked at the land about him he saw nothing to suggest that the shape of the earth is spherical. To him the earth was flat and mountains were merely bulges on a plain. How could he account for the appearance of rugged country as he approached it from the sea? First, he would see the highest peaks, then, as he got nearer, the lower foothills, and, finally, the beaches. He would have found it difficult to account for these observations if he believed the earth was flat, but quite easy if he assumed he was sailing over a curved surface. Although the Greek mathematician, Pythagoras (about 550 B.C.), taught his students that the earth is spherical in shape, this important fact was lost sight of for many centuries, and belief in a flat earth persisted through the Dark Ages.

It was not until the 15th century that men began to accept the evidence for a spherical earth. Columbus was one of the first to put the new idea to the test. He argued that, if the earth is round, the Indies could be reached by sailing west as well as east. The Indies, or East Indies as they are now called, were already well known to many adventurous Europeans. They were a source of riches—silks, spices, and precious stones. But the journey to the Indies was hazardous and took many weeks. To reach them traders sailed through the Mediterranean, then traveled overland hundreds of miles along a caravan route infested by brigands and, finally, by ship through the Indian Ocean. In 1492 Columbus set sail for the Indies by sailing west across the Atlantic. Eventually he reached a group of islands that he thought were the Indies. Today these islands are called the West Indies, reminding us of Columbus' attempt to reach the East Indies by sailing west.

A few years later, an expedition led by Magellan, a Portuguese navigator, sailed completely round the world. Leaving Seville, Spain, in 1519 the battered expedition returned to the same port from the opposite direction three years later. The voyages of Columbus and Magellan proved beyond doubt that the earth is round, not flat.

The Earth Rotates. The earth, like all other large heavenly bodies, rotates on its axis. How do we know this?

The sun appears to rise in the east, move across the sky from east to west, and set in the west. During the night it seems to continue its journey, returning to its starting point after 24 hours. But does the sun really move? Actually, of course, the earth makes a complete rotation on its axis every 24-hour day. In-

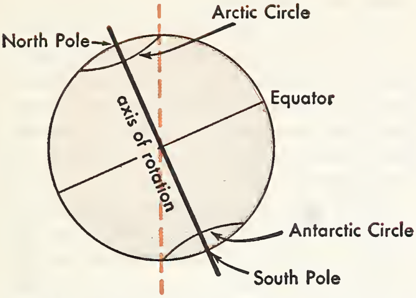


Fig. 4-1. The earth rotates about an inclined axis.

deed, while you have been reading this paragraph you have traveled many miles through space on the earth's surface.

The sun's rays illuminate one half of the earth at a time. If the earth did not rotate, one half of it would be in perpetual darkness and the earth would not be a pleasant habitation for man. But since it rotates, night succeeds day; or in other words, ***the rotation of the earth causes night and day.***

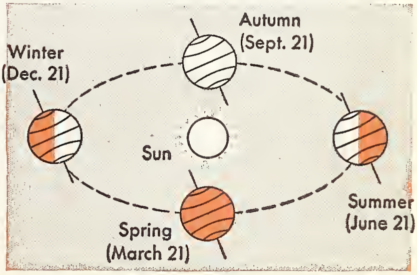
Notice again that ***a complete rotation constitutes one day.*** This is important since the day is a unit of time. The earth not only rotates on its axis, but it also revolves about the sun in an almost circular path or orbit. ***The revolution of the earth about the sun constitutes one year.***

The Earth's Axis Is Inclined. The earth's axis is not perpendicular to its orbit; it is tilted or inclined as shown in Fig. 4-1. Moreover, it always remains tilted in the same direction as it travels about the sun. What would happen if the axis were not inclined? In such a case all parts of the earth would be in sunlight for 12 hours, then in darkness for 12 hours for every day of the year. In other words, there would be no difference in the length of night and day throughout the year. Nor would there be any difference between summer and winter; that is, there would be no seasons.

The Four Seasons. Let us now see how the tilting of the earth's axis causes the different seasons. To get a clear idea of the relation between earth and sun, take a tennis ball to represent the earth, and push a knitting needle through it to represent the axis. Tilt the axis at an angle of about 25°. A light bulb placed at the center of a horizontally supported circular wire frame can be used to represent the sun and the earth's orbit. Now slowly rotate the tilted ball and, at the same time, move it along the wire orbit. Are there any positions on the orbit where the region about the poles is in darkness for a complete rotation? Are there any positions where the region about the poles is illuminated for a complete rotation? And are there any positions on the orbit where both poles are illuminated at the same time?

Let us now turn to the diagram (Fig. 4-2) which shows the positions of the earth in its orbit during spring, summer, autumn, and winter. Starting with the position marked ***summer,*** notice that the North Pole is ***toward*** the sun. Therefore, the region about the North Pole remains in the light of the sun throughout the entire 24-hour rotation. On the other hand, the South Pole is ***away*** from the sun and, therefore, is in darkness throughout the same 24-hour rotation. Does the same amount of light fall on the Northern and Southern hemispheres while the sun is in this position? It should be evident that all places north of the Equator receive light for more than 12 hours, and places south of the Equator receive light for less than 12 hours. Thus, in the Northern Hemisphere the days are longer than the nights, and it is summer; in the

Fig. 4-2. The four seasons.



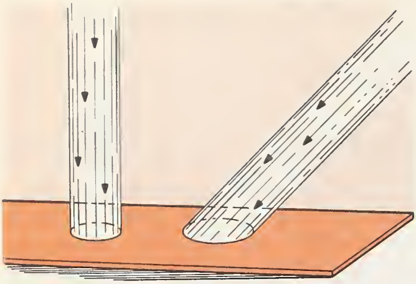


Fig. 4-3. The heat rays in the vertical beam are concentrated into a smaller area than the rays in the slanted beam.

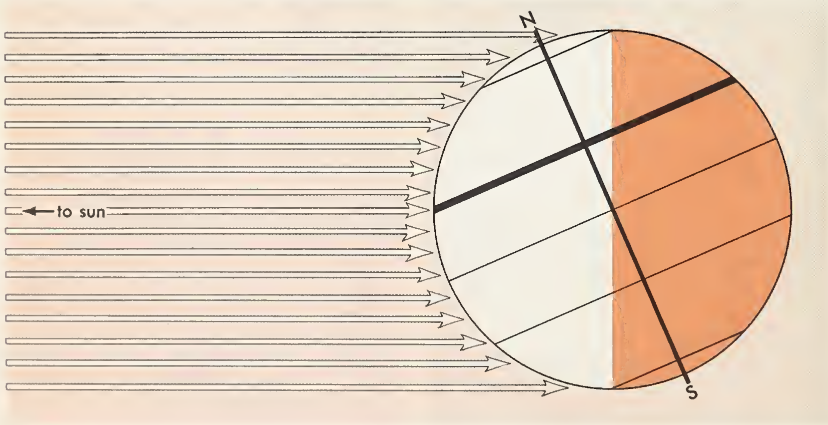
Southern Hemisphere, the nights are longer than the days, and it is winter.

Does the amount of heat as well as light vary in the two hemispheres? We must bear in mind that the sun supplies both heat and light. Moreover, the amount of heat that falls on the earth's surface depends on the slope of the sun's rays. Let us see why this is so. Examine the diagram (Fig. 4-3) which shows the two similar beams of light (that is, two circular beams of the same diameter). One beam strikes the earth vertically, and the other at an angle. Which beam will give the greater amount of heat over the same area of surface? The vertical beam forms a circular patch on the earth, whereas the slanting beam forms a larger oval or elliptical patch. The vertical rays therefore concentrate the heat of the sun over the smaller area and, as a result, they heat the ground more than the slanting rays.

With this in mind let us turn to the diagram (Fig. 4-4). The sun's rays are almost vertical at many places in the Northern Hemisphere, and at the corresponding places in the Southern Hemisphere, they are much more inclined. It follows that, under these conditions, the Northern Hemisphere is hotter than the Southern Hemisphere. ***This difference in the amount of heat and light received by the earth causes it to be warm in summer and cold in winter.***

Let us now turn to the position marked ***autumn*** in Fig. 4-2. Neither the north nor south pole is tilted toward or away from the sun so that all places in both hemispheres are illuminated for the same length of time. At the poles, at the equator, and at all places between them, there are 12 hours of darkness and 12 hours of daylight. This is the position of the autumn ***equinox;*** its actual date is Septem-

Fig. 4-4. In summer the sun's rays over the Southern Hemisphere slope more than the rays over the Northern Hemisphere. (The sun is overhead along the latitude, shown as a heavy wide line, north of the Equator.)



ber 21. (The word **equinox** is a Latin word meaning ***equal night.*** It means that the night (nox) equals the day.)

Now observe the positions marked ***winter*** and ***spring*** in the same diagram and see if you can answer the following questions. Why in December is the Northern Hemisphere colder than the Southern Hemisphere? Why is December 21, the shortest day in the Northern Hemisphere? Why is March 21 an equinox?

Finally, we must bear in mind that if the earth's axis were not inclined there would still be differences in ***climate,*** although no seasons. The poles would be much colder than the Equator, because the sun's rays would have the greater slant at the poles. However, because of the seasons and the varying length of night and day, the effect of the inclined axis is to intensify these differences in climate.

Latitude and Longitude. Suppose we wish to locate some place on the earth's surface, a town or a ship at sea, for example. How could we do it? First of all, we need points of reference, for without these the position of a body cannot be described. Suppose, for example, that someone asks you where a certain farm is. You might say, "Go down this road for two miles, then turn east along a dirt track for almost a mile." Notice that these instructions include two points of reference (the starting position and the near end of the dirt track) and also two distances in different directions.

Precisely the same steps must be taken if we wish to locate a town on a map or a ship at sea. However, in this case, the reference "points" are lines, or rather circles, and distances are expressed in degrees, not miles. Measurements expressed in this way are called ***latitude*** and ***longitude.***

Latitude. The line of reference for north-south measurements is the Equator. ***Latitude is, in fact, distance north or south of the Equator,*** and parallels of latitude are imaginary circles that run around the earth parallel to the Equator. It is apparent that the parallels of latitude decrease in size as they approach the poles. The largest is at the Equator and the smallest, which is reduced in size to a point, is at either pole.

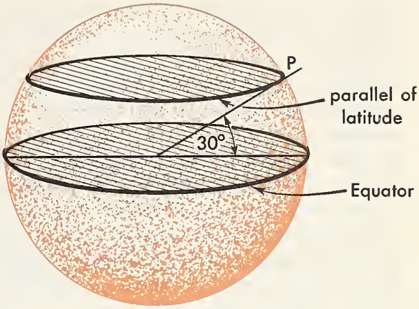


Fig. 4-5. What is meant by a latitude of 30° N?

What is meant by a latitude of, say, 30°? First, at the centre of a circle representing the earth we will make an angle of 30° with the plane of the Equator. The arm of this angle (see Fig. 4-5) strikes the surface of the earth at a point ***P.*** Now we will draw a circle around the earth through ***P*** parallel to the Equator. This circle is the thirtieth parallel of latitude. Notice that all points on the same parallel of latitude are the same distance from the Equator. In the same way, we could draw parallels of latitude any number of degrees apart; for example, 1°, 5°, 10°, 30°, 60°, and so on. It should now be clear that the latitude of the Equator is 0° and the latitude of each pole is 90°.

One degree of latitude measures about 70 miles. This distance is found by dividing the circumference of the earth (24,800 miles) by 360 (the number of degrees in a circle). Ottawa is almost on the 45th parallel of latitude in the Northern Hemisphere, or, as we say, 45° N. Its distance from the Equator is, therefore, about 3200 miles. (How is this distance computed?)

If we wish to be more exact we express latitude in degrees and minutes; 60 minutes make one degree. A more exact latitude of Ottawa, using symbols, is 45°24′ N. The latitude of London is 51°28′N; and that of Melbourne (Australia) 37°55′S.

Longitude. The line of reference for east-west measurements is the great circle that passes through the poles and London, England. Actually, the place of reference is the Royal Observatory near London, a reference point adopted by practically all countries. The longitude of this particular starting point is, like the equator, zero or 0°.

To draw other lines of longitude we must divide the circle of the Equator into 360°. If we now draw circles through the North and South poles and through each of these points on the Equator we would have 180 circles of longitude. It follows that circles of longitude are all of the same size, the size of the Equator. For this reason, they are referred to as ***great circles.*** As Fig. 4-6 shows, they resemble the divisions of an orange being furthest apart at the Equator and converging at the poles.

Longitude is measured both east and west from London, 180° to the east and 180° to the west, making a total of 360° in all. The longitude of Ottawa is about 75° W, that is about 75° west of London; the longitude of Bombay is about 75° E, or about 75° east of London. Places located on longitude 180° are on the other half of the great circle that passes through London; that is, on the opposite side of the earth.

In this respect the measurement of longitude differs from latitude. Every great circle of longitude constitutes two lines of longitude, or *meridians.* The near half from pole to pole is one meridian, and the further half from pole to pole is the other. The further half of the meridian through Ottawa passes through Singapore and its longitude is roughly 105° E. (Notice that 75° W + 105° E = 180°.) Therefore, Singapore is almost exactly on the opposite side of the earth from Ottawa.

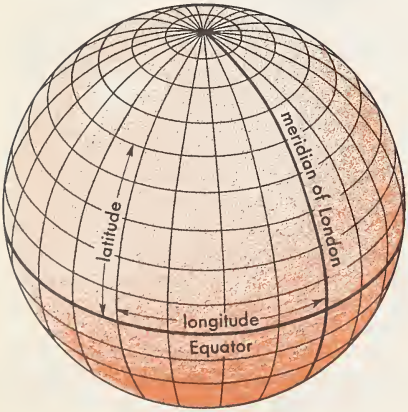
As in the case of latitude, minutes as well as degrees are used for more exact measurements of longitude. Thus the following cities can be accurately located on the globe from the following information:

Ottawa	lat. 45°24′N,	long. 75°43′W
London	lat. 51°28′N,	long. 0°0′
Melbourne	lat. 37°55′S,	long. 145°0′E

Time. By general agreement the day starts at midnight, or zero hours. The middle of the day is 12 hours or 12 o'clock noon; the hours between midnight and noon are referred to as A.M. (ante meridiem, "before noon"); the hours between noon and midnight are called P.M. (post meridiem, "after noon").

To avoid confusion, a 24-hour time-day is gradually being adopted. It is already used in the armed forces, by astronomers and in some European countries. In this new system all A.M. times are the same as our present A.M. times, but the P.M. times are expressed in numbers ranging from 13 to 24. For instance, 4 P.M. becomes 16 hours, and 9:30 P.M. becomes 21:30. The advantage of this system is that there is never any doubt

Fig. 4-6. Latitude and longitude. Notice that all lines of longitude are great circles, whereas only one of latitude (the Equator) is a great circle.



whether the time referred to is morning or evening.

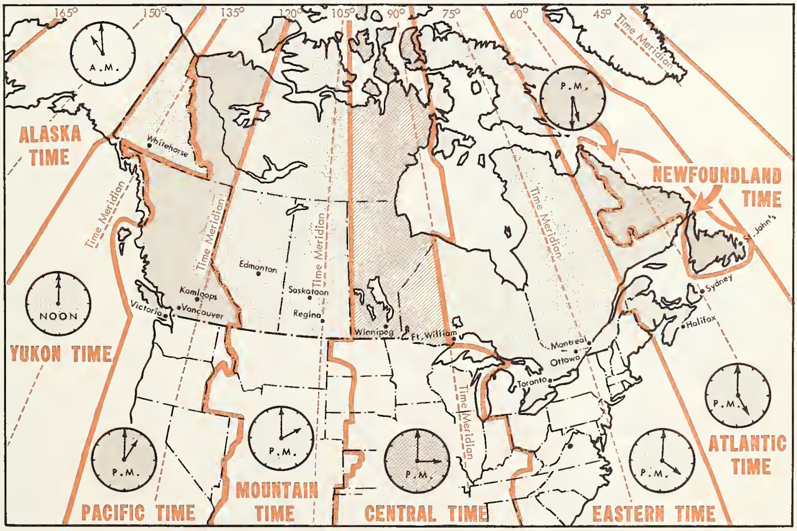
The words ***noon*** and ***midday*** have been used to represent a particular time of day. How is this particular time determined? Midday is the time when the sun is highest in the sky, or at its zenith as astronomers say. Another glance at Fig. 4-6 will convince you that ***places on the same longitude have the same time.*** When it is midday in Ottawa it is midday at all places on longitude 75°43′W; that is, at all places north and south of Ottawa. It is for this reason that lines of longitude are called meridians, a word derived from Latin meaning "middle of the day."

Time Belts. As already stated, midday is the time when the sun is highest in the sky. It follows that the midday time in your hometown is different from the midday time of nearby towns.

In the early days, towns had their own local time which resulted in much confusion, particularly when railroad companies tried to make up time tables. To avoid the confusion the country was divided into time zones, so that all clocks in the same zone would be at the same time. The system of international standard time was adopted in 1884 largely as a result of the efforts of Sir Sandford Fleming, a Scotsman, who also had much to do with the building of the Canadian Pacific Railway across Canada.

What is a time zone? The earth makes a complete revolution through 360° in a day of 24 hours; that is, it rotates through $\frac{360^\circ}{24} = 15^\circ$ in one hour. Hence 15 degrees of longitude represents a time zone of one hour and the globe is divided into 24 zones each 15° wide. Let us apply this to Canada. The distance from the east to the west coast is roughly 88° of longitude, a distance equal to six zones. These six zones are based on the sun time on the 60th, 75th, 90th, 105th, 120th, and 135th meridians, which are approximately the longitudes of Sydney, Montreal, Fort William, Saskatoon, Kamloops, and Whitehorse. You will notice from Fig. 4-7 that there are actually seven rather

Fig. 4-7. Time belts in Canada and the United States.



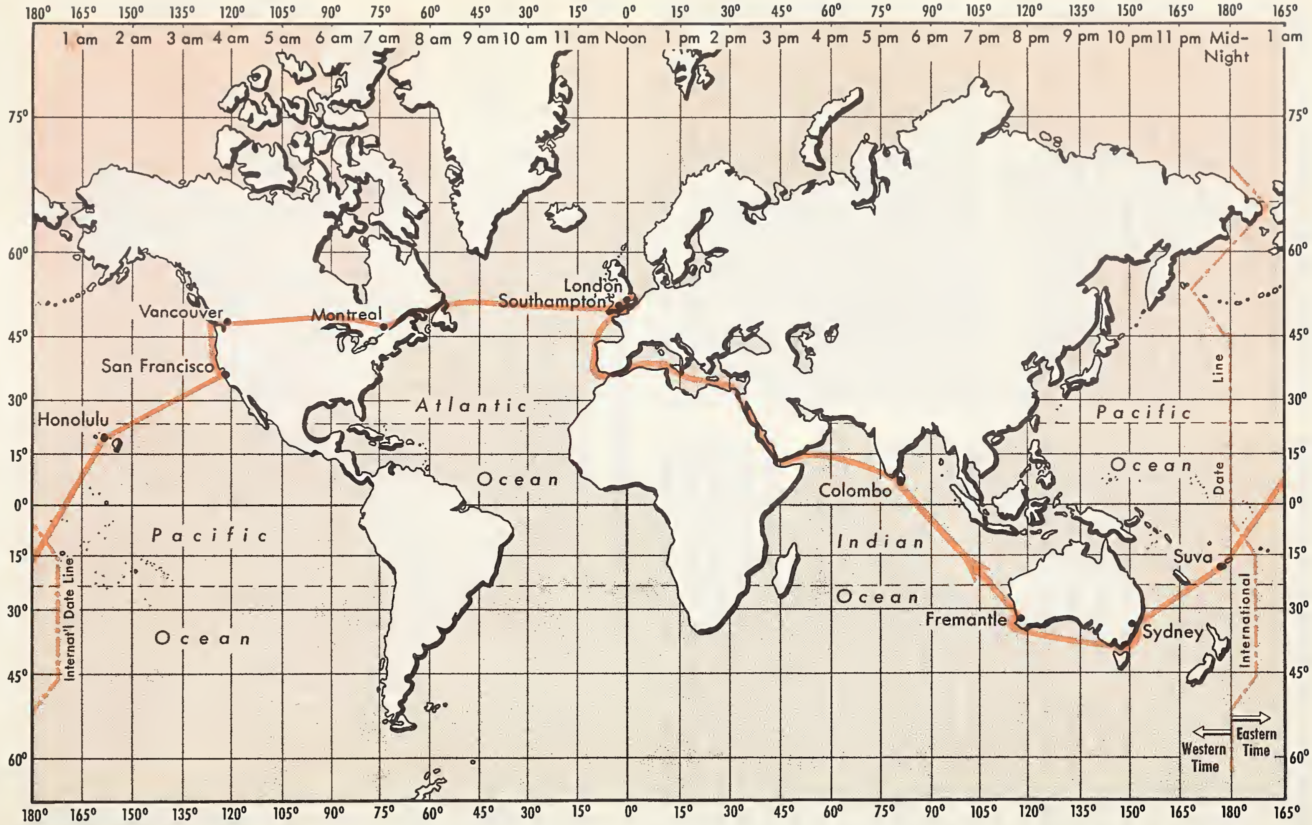


Fig. 4-8. In a trip around the world clocks are adjusted 24 times.

than six different times in Canada, the seventh being Newfoundland Time, which is half an hour ahead of the time in Halifax. What reasons can you give for this difference in time between Newfoundland and Nova Scotia?

The seven time zones are known as Newfoundland Time, Atlantic Time, Eastern Time, Central Time, Mountain Time, Pacific Time, and Yukon Time. The zones, as the map shows, are not necessarily bounded by the meridians. Deviations are made so that inhabitants in a thickly populated region will not be inconvenienced by two different standard times in the same locality.

As you probably know from experience, time zones offer no difficulty to the traveler. If you travel west from Halifax to Whitehorse, you will have to adjust your watch every time you pass from one time zone to the next. Will your watch be set forward or back? By how much? If you travel east from Victoria to Toronto what adjustment will you make in your watch?

International Date Line. Since there are 24 time zones across the surface of the globe, each time zone represents 1/24 of the earth's circumference of 360°. If we make a trip around the world, clocks will therefore have to be adjusted 24 times by the time we return to our starting point. This will result in losing a day or gaining a day depending on whether clocks are put back or forward. But let us investigate the problem by taking an imaginary trip round the world, and following the route on the map (Fig. 4-8).

Let us start from Montreal and travel westward. First we will take the train to Vancouver. Watches will, of course, be set back three hours on this part of the trip. At Vancouver we will board a ship that takes us to Sydney, Australia. At sea, clocks are usually adjusted at midnight so that the ship's time will register noon when the sun is at its zenith on the following day. Our first port overseas is Honolulu in the Hawaiian Islands. The longitude of Vancouver is about 125°W and that of Honolulu is 155°W. Therefore, clocks are retarded two hours. Why? Our next port of call is Suva in the Fiji Islands. The longitude of Suva is about 180°W so that clocks are again retarded 2 hours on the trip to Suva. We now continue the voyage to Sydney on the south east coast of Australia, a city whose longitude is about 150° E. Again the clocks are retarded 2 hours. At Sydney we disembark and board another boat for London. So far, since leaving Montreal, the clocks have been retarded nine hours. In other words, if it is midday in Montreal, our clocks in Sydney register, at the same instant, 3 A.M. But let us continue the journey.

The ship from Sydney follows the south coast of Australia. It calls at Melbourne and then at Fremantle, near Perth, on the southwestern tip of Australia. The longitude of Perth is about 115° E and so clocks are retarded 2 hours from Sydney time. From Fremantle the ship sails north west through the Indian Ocean to Colombo in Ceylon, whose longitude is about 80° E. On this part of the trip the clocks are retarded 2½ hours. Why? From Colombo the ship passes through the Red Sea, the Suez Canal, the Mediterranean Sea, and on to London. The longitude of London is, of course, 0 ° so that the clocks are again retarded 5½ hours after leaving Colombo. Explain why. At London we must change ships. So far, however, since leaving Montreal, clocks have been retarded 19 hours. If, at this instant, it is midday in Montreal, the clocks in London register 5 P.M. on the previous day!

From London we take a train to nearby Southampton, board another ship and set sail for Montreal. In this ocean trip we cover about 75 degrees of longitude so that clocks are again retarded 5 hours. We have now completed our trip round the world and our watch has been put back 24 hours, or one whole day. This means that if Montreal time is midday Wednesday when we arrive, it is, by our reckoning, midday Tuesday; we have

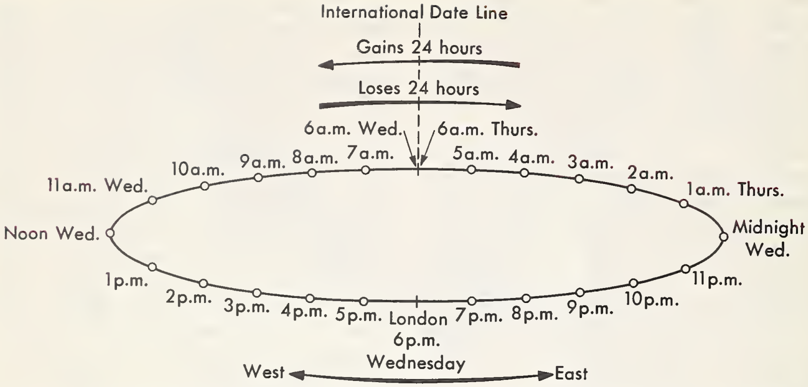
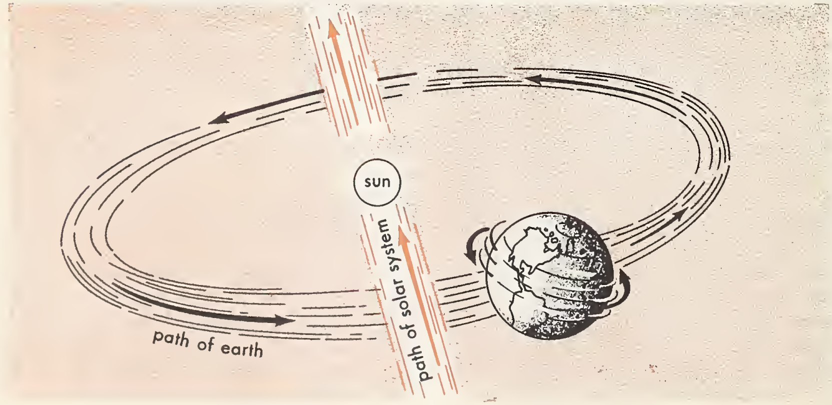


Fig. 4-9. A west-bound ship jumps a day when it crosses the international date line; an east-bound ship gains a day.

"lost" a whole day. If we had traveled eastward round the world instead of westward we would have gained a whole day and, by our reckoning we would have arrived in Montreal at midday Thursday instead of midday Wednesday.

How can this confusion be avoided? Clearly a time adjustment of one day is necessary. By international agreement, this adjustment is made when ships cross the 180th meridian, known as the ***international date line***. Just west of this line, the time is known as Western Time; just east of it, the time is known as Eastern Time. Western and eastern times differ by one whole day. When a west-bound ship crosses the international date line it jumps a day to get into the Eastern Time zone. If you happen to cross the line about Tuesday midnight, the following morning will be Thursday; Wednesday has disappeared. On the other hand, if the date line is crossed by an east-bound ship it "gains" a day. If the line is crossed about midnight on Tuesday, the following morning will be Tuesday, so that, in

Fig. 4-10. The earth has several different motions: it rotates on its axis; it revolves around the sun; and as part of the Milky Way, it moves about the center of the galaxy.



that particular week you will have an extra Tuesday. Of course, the day "lost" or "gained" is made up by setting our watches forward or backward as we travel round the world.

Astronomical Time Clocks. Have you ever wondered why there are 365¼ days in a year? Why this awkward number which means that we have to have a leap year every fourth year? The difficulty arises because in measuring a day and a year we use two entirely different "clocks" for measuring time.

As already stated, the rotation of the earth on its axis constitutes a day, and the revolution of the earth around the sun constitutes a year. If, therefore, we express a year in terms of days we cannot expect to be lucky enough to get a convenient whole number. But the situation is further complicated by introducing still a third time clock. This is the moon's clock of one month, the time taken for the moon to revolve about the earth. Expressed in days, it is almost 29 days. A week is a quarter (or 90 degrees) of a rotation of the moon around the earth. Our calendar attempts the impossible task of coordinating the three movements of earth, moon, and sun in relation to each other. No wonder the calendar lacks precision and is such an unsatisfactory time table.

The Earth's Motions. So far we have considered two different motions of the earth, its rotation about the axis and its revolution about the sun. Does the earth have other motions? You may recall that the earth has at least four different motions: (1) as a planet, (2) as a member of the solar system, (3) as part of the Milky Way galaxy, (4) as part of the expanding universe. It spins about its axis once a day at 1000 miles an hour (or 0.28 miles a second) for places on the Equator; it whirls around the sun at 20 miles a second; it moves around the center of the Milky Way at 140 miles a second; and it flies through space at thousands of miles per second as the universe expands. This is a far cry from the flat and stationary earth of the ancients.

Things to Remember

The earth makes ***one complete rotation*** on its axis every 24 hours.

The rotation of the earth causes ***night and day.***

The revolution of the earth about the sun constitutes ***one year.***

The ***four seasons*** are caused by the tilting of the earth's axis and the motion of the earth around the sun.

The ***spring equinox*** is March 21, and the ***autumn equinox*** is September 21.

Latitude is distance north or south of the Equator.

Longitude is distance east or west of London (England).

Places in the same longitude have the ***same time.***

There are ***seven time zones*** in Canada: the Newfoundland, Atlantic, Eastern, Central, Mountain, Pacific, and Yukon time zones.

The ***international date line*** is longitude 180°. Ships sailing west set their calendars ahead one day on crossing the line.

Questions

1. Explain the meanings of the words or terms: (*a*) equinox, (*b*) meridian, (*c*) P.M.

2. What single fact do you associate with each of the names: (*a*) Columbus, (*b*) Magellan, (*c*) Pythagoras.
3. Explain the significance of the term ***spring equinox.***
4. The latitude of Winnipeg is 49°54′N and its longitude is 97°8′W. What does this mean?

GROUP B

5. (*a*) The earth has several kinds of motion. Name them. (*b*) Which of these motions are used to measure time? Explain.
6. In the month of June the days are longer than the nights in the Northern Hemisphere. Explain why.
7. In the month of June it is warmer in the Northern than in the Southern Hemisphere. Explain why.
8. The circumference of the earth is nearly 25,000 miles. (*a*) Calculate the distance in miles represented by 1° latitude. (*b*) The latitude of Capetown is 35° S. How far is Capetown south of the Equator?
9. "The earth rotates on an inclined axis." Explain the effects of (*a*) the rotation of the earth, and (*b*) the inclination of its axis.
10. (*a*) What is a time zone? (*b*) How is it computed? (*c*) Why are time zones used in Canada?
11. Explain why there is a difference of four hours in time between Halifax and Vancouver. Which of these cities is ahead in time? Why?
12. What is the international date line? Why is this date line necessary?
13. If it is 2 P.M. Wednesday in Calgary (long. 114° W), what time and day is it in (*a*) Fairbanks, Alaska (long. 145° W), (*b*) Tokyo, Japan (long. 140° E), (*c*) London, England (long. 0°), (*d*) Auckland, New Zealand (long. 180°).

Things to Do

A shadow cast by the sun · Set up a pole in a piece of level ground and observe the shadow cast by the sun. Place a marker at the end of the shadow every hour for several hours during the same day. Measure the angles between the lines running from the pole to the markers. These angles should be 15°. Why?

Notice that the ***direction*** of the shadow changes. In the morning it points west; in the evening it points east. What does this indicate?

Also observe the ***length*** of the shadow. It is long in the morning and evening and shortest at noon. What does this indicate?

Make scale drawings of the length of the pole and the lengths of the shadows at such hours as 8 A.M., 10 A.M., 12 noon, 2 P.M., 4 P.M., and 6 P.M. Measure the angles of inclination of the sun's rays: What do these drawings indicate?

If this experiment is repeated at different times of the year you will observe that the inclination of the sun's rays depends on the particular season.

5

ROCKS OF THE EARTH (1)

MINERALS AND IGNEOUS ROCKS

THE CRUST of the earth is a rigid shell only about 30 miles thick, less than one hundredth of the distance to the center. Only a small fraction of the crust is exposed to view so that much of our knowledge of it comes from indirect observations. We do know, however, that the deeper we penetrate into the crust the hotter it becomes. This is proved by taking temperatures in deep mines and well borings. In searching for oil some wells have been drilled to a depth of more than 20,000 feet, or almost 4 miles.

The average rise in temperature is about 30 degrees centigrade for every mile of descent. At a depth of about three miles the temperature of boiling water is reached, and at a depth of 30 miles the temperature is high enough to melt rocks. This is where the lavas of volcanic rocks are usually found. We must not think, however, that all rocks are molten at this depth for, as will be explained in Chapter 8, most rocks under the crust are in solid condition.

Geysers are, of course, direct evidence of heat within the crust. They are common in regions of volcanic activity, in Italy, Japan, New Zealand, and western U.S.A. In such places molten rock, from lower levels, comes near the surface. If water seeps through cracks in the crust, it is heated by contact with the hot rocks below. The water boils, and the pressure of the steam ejects some of the hot water. Old Faithful, the geyser in Yellowstone National Park, is a familiar example of this process. In some of the geysers in Italy, Japan, and New Zealand, high temperature steam is ejected, "dry" steam as it is called. This is used as a source of energy, providing electric power for cities and for industry.

Minerals. All rocks (for example, granite, basalt, limestone) are made of minerals. What then is a mineral? A mineral is a combination of substances; it always has the same chemical composition and its own particular crystalline structure. Quartz, feldspar, mica and calcite are well-known minerals. Can you name others? A rock may contain two, three, four, or even more different minerals. The relation between a rock and its minerals is similar to the relation between a fruit cake and its ingredients. If the rock corresponds to the cake, the minerals correspond to the raisins, nuts, cherries, and lemon rind.

What are minerals made of? It is evi-

Fig. 5-1. Old Faithful, the well-known geyser in Yellowstone Park.

dent that minerals are made of elements, but relatively few of them are present in large amounts. Only eight elements are important in the composition of most of the common minerals. The following is a list of these elements with a percentage figure indicating their abundance in the earth's crust.

Element	*Symbol*	*Percentage by Weight*
Oxygen	O	46.7%
Silicon	Si	27.7%
Aluminum	Al	8.1%
Iron	Fe	5.1%
Calcium	Ca	3.7%
Sodium	Na	2.8%
Potassium	K	2.6%
Magnesium	Mg	2.1%
		98.8%

Notice that these eight elements account for almost 99 per cent of the earth's crust. The most abundant is oxygen. But, of course, since it is a gas, oxygen never occurs free in the crust; it is always combined with other elements. The next most abundant element is silicon. Oxygen and silicon make up almost 74 per cent of the crust. It is evident that most minerals must contain the elements oxygen and silicon in their make-up. They are called ***silicates***. In other words, silicates always contain the elements oxygen and silicon, and, in addition, they usually contain at least two other elements. We will now consider a few minerals, and from the long list we shall select only feldspar, mica, quartz, olivine and calcite.

Feldspar. Feldspar is, by far, the most plentiful mineral; it is a constituent of all granites. Its name is derived from a German word which means "field mineral," a name that indicates its abundance. There are several kinds of feldspars but all are aluminum silicates; that is, they all contain aluminum, silicon, and oxygen plus another element. It is this other element that enables us to classify the feldspars.

All the feldspars have a ***cleavage,*** that is a tendency to split easily along a surface. Indeed, they have two cleavages and split in two directions either at right angles or nearly at right angles to each other. They show a variety of colors, the most common being white, gray, or pink.

The ***specific gravity*** of the feldspars is

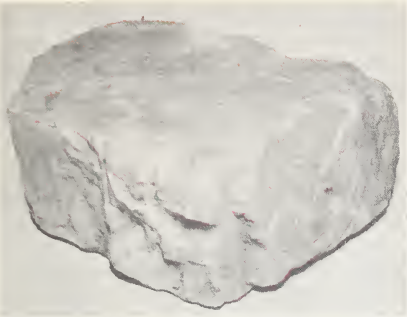
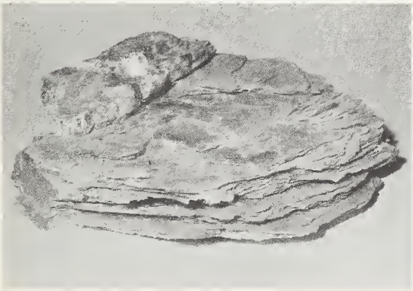


Fig. 5-2. A specimen of orthoclase feldspar. The rock shows the two cleavage directions almost at right angles to each other.

about 2.7. This means that ***a given volume of feldspar is 2.7 times heavier than the same volume of water.*** It is important to remember the meaning of the term specific gravity since we shall make frequent reference to it. If the specific gravity of a rock or mineral is less than 3.0 it is called a ***light*** rock; if its specific gravity is greater than 3.0, it is called a ***heavy*** rock.

The two common varieties of feldspar are ***orthoclase*** and ***plagioclase.*** Orthoclase is so named because its cleavages are exactly at right angles to each other (the Greek word ***orthos*** means ***straight,*** and ***klasis*** means ***a break***). In orthoclase, the element potassium is attached to the alumino-silicate group, and its formula is $KAlSi_3O_8$. Orthoclase is used extensively in making china.

Fig. 5-3. Mica is a soft, shiny mineral that easily flakes into thin sheets.



Mica. Mica, like feldspar, is also an aluminum silicate. It is a shiny mineral which cleaves into thin sheets very easily. Also like feldspar its name is a descriptive one, being derived from the Latin word ***micare,*** to shine. Mica is a little heavier than feldspar; its specific gravity is about 3.0.

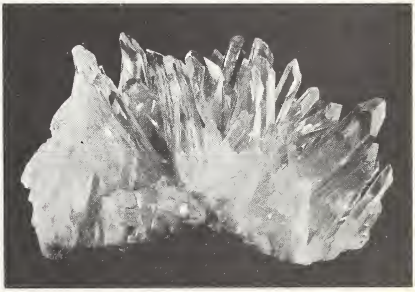
There are two varieties of mica, white mica or muscovite, and black mica or biotite. Muscovite is the best electrical insulator and is used in electrical equipment such as generators, condensers and telephones. It is also used as a framework to support heating units of electric toasters, and as windows in the doors of steel furnaces. Indeed, the name muscovite is indirectly derived from its property of transmitting light. In old Russia, known as Muscovy, this mineral was commonly used by the Muscovites as a substitute for glass in their crude houses.

Muscovite and biotite are, like the feldspars, found free in nature. They are also a common constituent of granites. Examine a granite which contains mica and notice that this mineral can be peeled off with the point of a knife.

Quartz. Quartz is the common name for the mineral ***silica*** or silicon dioxide, SiO_2. From a glance at the table on page 43, it is not surprising that quartz is the most plentiful oxide on earth; as a rock-former it is second only to feldspar in abundance. Its specific gravity is 2.7, the same as that of feldspar.

When free to grow, quartz forms beau-

Fig. 5-4. Quartz crystals.



tiful glassy crystals, hexagonal in shape, with a pyramid at each end. In nature, however, there is usually some interference with growth. As a result, a cluster of quartz crystals is usually found attached to a common base so that the pyramid is developed only at the free end. However, in most rocks, for reasons given later, quartz has no definite shape.

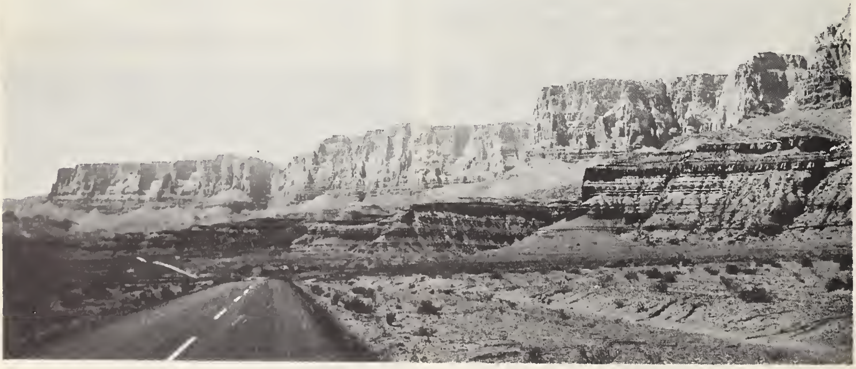
Quartz is not very soluble in water. But under the right conditions, the small amount of it dissolved in sea water will replace wood in the trees of submerged forests. In this way, the well-known petrified forests of Arizona have been formed. Quartz deposited in this way is usually called agate.

Quartz is the hardest of the common minerals and is used as an abrasive as, for example, on sandpaper. If quartz crystals are colored by the right kind of impurity they are semiprecious stones, such as amethyst and opal. A most important use of quartz crystals is in radio and television transmitters. If a thin slab of quartz is compressed it becomes electrically charged, one side of the plate being positively charged and the other negatively. Conversely, if the plate is electrically charged it becomes distorted, an effect that causes it to vibrate thousands, or even millions, of times per second. Since a quartz plate vibrates at a constant rate, it is used to control the "frequency" of a radio station. As a result, the station can always be picked up at the same place on the dial of a radio receiver.

Olivine. The mineral olivine has a yellowish-green color, the color of olives, from which it is named. It is a silicate rock but, unlike feldspar and mica, it contains no aluminum. Instead, the silicate group is attached to atoms of magnesium, or iron, or both and, for this reason, it is called a ferromagnesium silicate. Olivine is one of the heaviest minerals; its specific gravity is about 4.0. You will recall that geologists think there is a stony shell of heavy rock, probably olivine, below the basalt layer of the crust (see page 6). Does the above evidence support this assumption?

Calcite. Calcite is one of the few minerals that does not contain silica. It therefore cannot be a silicate. Instead it is a carbonate; its chemical formula is $CaCO_3$. What elements are present in its makeup? And what is its ***chemical*** name? Calcite is a light mineral; its specific gravity is 2.7. It is the chief mineral in the rocks limestone, chalk, and marble. Usually it is white in color but may be tinted red, as in the colorful red limestones of Bryce Canyon.

Fig. 5-5. Stratified rocks in Vermillion Cliffs, Kaibab National Forest, Arizona. Notice that the sedimentary rocks are arranged in horizontal layers.



Limestone, chalk and marble are common rocks in the earth's crust and the geologist has a simple test by which he can recognize them, the so-called acid test.

Demonstration

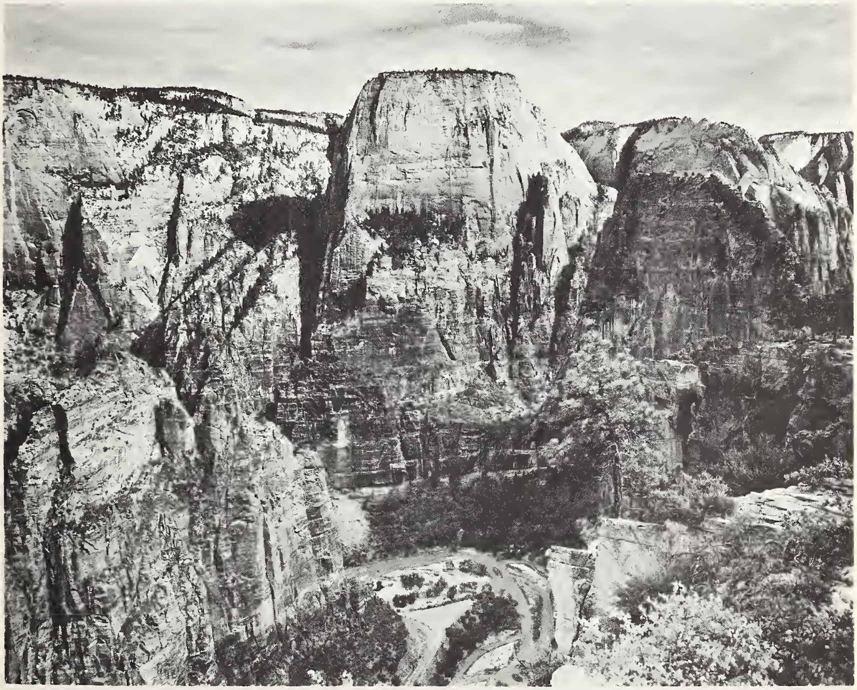
5-1. A chemical test for calcite

Using a glass tube, place a few drops of dilute hydrochloric acid on calcite (or limestone, chalk or marble). There is a brisk effervescence and a gas (carbon dioxide) is evolved (see page 176). Now try the same test on quartz, olivine, and feldspar. No effervescence occurs in these cases. Why not?

Rocks. Rocks may be studied wherever they crop out on the surface. Such outcrops can be seen in quarries, in gorges carved out by rivers, in mountain cliffs, or even in an open field if the topsoil has been washed away. Some outcrops are bedded, that is, they consist of layers or strata as in the Rocky Mountains; elsewhere they are uniform or unstratified, as in the peaks of Zion Canyon. These two types give rise to the classification of rocks into two main groups: the stratified rocks are called ***sedimentaries,*** the unstratified rocks are called ***igneous.*** A third group, derived from igneous or sedimentary rocks is known as ***metamorphic*** rock.

Igneous Rocks. Igneous rocks are those which have crystallized from the molten condition. (The word ***igneous*** is derived from the Latin word ***ignis,*** meaning fire.) They are the original rocks of the earth, the ancestors of all sedimentary and meta-

Fig. 5-6. Unstratified or igneous rock. Zion Canyon, Utah.



morphic rocks. It should not, therefore, surprise you to know that igneous rocks are the predominant type. Indeed, 95 per cent of the rocks of the crust to a depth of ten miles are igneous, and only 5 per cent are sedimentary and metamorphic.

Molten rock in the earth's interior is called ***magma;*** on the surface it is called ***lava.*** Both magma and lava cool and solidify to form rocks of similar composition. But the nature of these rocks varies considerably.

Why should magma and lava form different rock types? At great depths below the surface magma cools and crystallizes very slowly. This slow cooling process gives the molecules of the minerals ample time to align themselves into the crystal structure. The result is that the crystals are large and easily visible to the naked eye. Rocks of this kind are called ***coarse-grained.*** Granite is an example of a coarse-grained rock. Molten lava, however, is exposed to the atmosphere and it therefore cools rapidly. How does this affect the rock structure? The various rock molecules do not have time to align themselves to form large crystals and their crystals usually can only be seen with the aid of a microscope. Rocks of this kind are called ***fine-grained.*** Pumice is a fine-grained rock. Thus the main difference between granite and pumice is due entirely to the conditions under which they cooled and solidified.

Rocks melt at such high temperature that it is impossible to show rock crystallization as a simple laboratory experiment. However, the element sulfur melts and crystallizes at a temperature only a little above the boiling point of water, and is a convenient substitute for rock.

Demonstration

5-2. Crystallization of molten sulfur

Half fill a dry test tube with powdered sulfur. Hold the test tube in a clamp and heat it very gently in the hot gases above a Bunsen flame until it has all melted. Place a folded filter paper in a funnel on a ring stand and pour the molten sulfur into the filter. When a solid crust has nearly covered the surface, pour the remaining molten sulfur into a beaker of water. Break open the filter paper to expose the crystals. Notice the needle-shaped crystals that grew as the molten sulfur cooled and solidified. Notice also the sulfur in the beaker. This sulfur was cooled rapidly by the water so that no crystals can be observed in this solid mass.

Order of Crystallization. We must not think that all the minerals in a magma crystallize at the same time. Like sulfur, all minerals have their own particular melting point (or crystallizing temperature). Suppose a molten mixture contains biotite, feldspar and silica. Biotite is the first mineral to crystallize and then, at a lower temperature, feldspar crystallizes. Most of the silica is, of course, used up in the formations of biotite and feldspar. Any left over now crystallizes as quartz. Quartz, being the last to crystallize, is unable to develop its characteristic hexagonal shape; it merely fills in the spaces between the other crystals. That is why, in a granite rock, quartz is seen as glassy blobs packed in between the feldspar and other minerals.

Igneous rocks are divided into two large groups, intruded and extruded. An ***intruded*** rock is formed when a mass of magma invades the earth's crust and crystallizes within the crust. An ***extruded rock*** is formed if the magma escapes through the crust and crystallizes on the surface. Granite is the most abundant intruded rock, and basalt the most abundant extruded rock.

Granite. The uppermost 10 miles of the continental areas of the crust consist largely of granite. Outcrops of granite can be seen best in the Canadian Shield, in the White Mountains of New Hampshire (the Granite State), in the Black Hills of South Dakota, and in the Coast Range of British Columbia. It also lies under the prairies, where it is covered by sedimentary rock and topsoil.

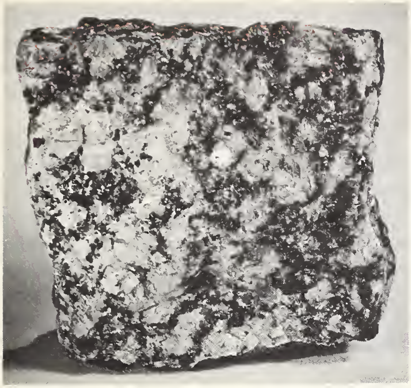


Fig. 5-7. Biotite granite. Notice the black crystals of biotite and the whitish crystals of feldspar.

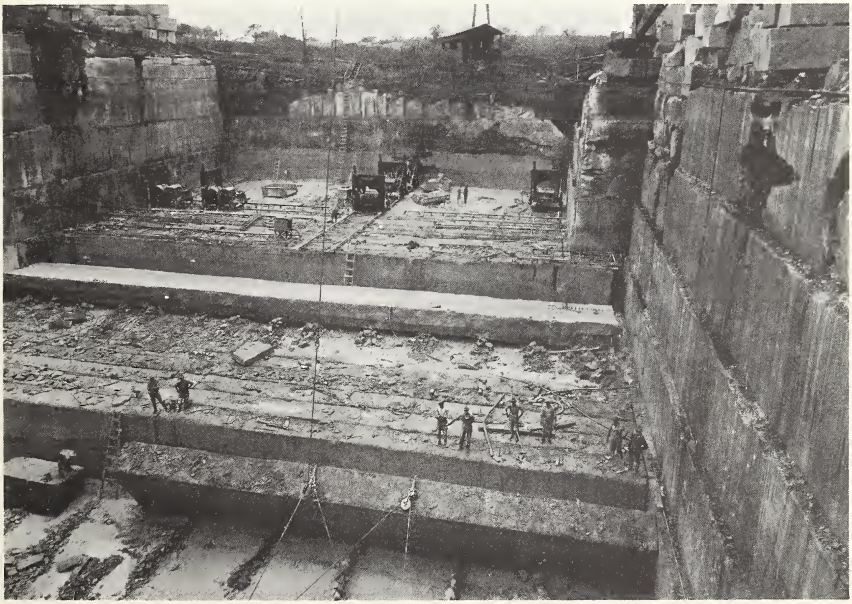
Granite crystallizes deep down in the earth's interior and is made up of at least three minerals. It always contains feldspar and quartz; the third mineral is usually mica (muscovite or biotite). The color of granite varies from gray to pink depending on the nature of the feldspar.

Granite is a durable strong rock that resists weathering. It is therefore used as foundations for massive buildings, as piers for bridges, and for sea walls. Moreover, it takes on a high polish, a property that accounts for its use in monuments and ornamental columns.

Basalt. Basalt is the most abundant fine-grained rock. It is a lava, a volcanic rock, and was ejected for the most part through cracks in the crust. In past ages, there were extensive flows of it and the rock now covers wide areas in many parts of the world. The volcanic deposits in the northwestern United States, particularly in Idaho, were formed in this way.

Basalt is a mixture of plagioclase feldspar and ferro-magnesian minerals. Unlike granite, it contains no free quartz. Also, because of the high proportion of iron that it contains, basalt is a heavier rock than granite and darker in color.

Fig. 5-8. A granite quarry. The immense pieces of hewn stone weigh many tons.



Pumice and obsidian are also volcanic rocks. Pumice is the name given to a spongy rock that solidified as steam and other gases bubbled through it. It is so porous that it will float on water. Artificial "pumice" can be made by blowing steam through molten slag. The product is called ***rock wool*** and is used for insulating houses.

Sometimes lava cools so rapidly that it is completely non-crystalline and looks like glass. Obsidian is a black lava formed in this way. However, obsidian is much harder than bottle glass and was used by the Indians in making arrow heads. Probably the best known obsidian deposit is Obsidian Cliff in Yellowstone National Park, Wyoming. This outcrop is about 9 miles long and 5 miles wide, rising 250 feet or so above the countryside.

Finally it is interesting to compare the weights of some of the igneous rocks. Granite (specific gravity about 2.7) is lighter than basalt (specific gravity about 3.0); basalt is lighter than olivine (specific gravity about 4.0); olivine is lighter than iron (specific gravity about 7.5). Do these data on specific gravities support the conclusions reached on page 6?

Things to Remember

Minerals are constituents of rocks; they are chemical compounds.

Oxygen and silicon, are the ***most abundant elements*** in the earth's crust.

Feldspar is the ***most abundant mineral.***

Specific gravity of substance $= \dfrac{\text{weight of substance}}{\text{weight of same volume of water}}$.

Quartz is the ***most abundant oxide*** in the crust.

Rocks are divided into two main groups: ***sedimentary*** (or stratified) and ***igneous*** (or unstratified).

Igneous rocks have crystallized from the molten condition. Ninety-five per cent of the rocks of the crust are igneous.

Granite is the most abundant intruded rock; ***basalt*** is the most abundant extruded rock.

Questions

GROUP A

1. (*a*) What is a mineral? (*b*) Name three minerals. (*c*) Name the two most plentiful elements usually found in minerals.
2. The specific gravity of basalt is 3.0. What does this mean?
3. State some of the industrial uses of (*a*) mica, (*b*) quartz.
4. What is an igneous rock? Give three examples.

GROUP B

5. Explain how a geyser is formed.
6. Describe the mineral feldspar.
7. (*a*) What are the constituents of granite? (*b*) How was granite formed from molten rock?
8. Explain the following terms and give examples: (*a*) coarse-grained rocks and fine-grained rocks, (*b*) intruded rocks and extruded rocks.

9. What is basalt? How does basalt differ from granite?
10. State the similarities and differences between pumice and obsidian.

Things to Do

Rock collection. Minerals and igneous rocks · If you are interested in geology you should collect rock specimens. Begin with such common rock-forming minerals as orthoclase feldspar, biotite, mica, and quartz. Your collection should also include several samples of granite and basalt. You should be able to identify the minerals in granite.

Identify and label each specimen and, if you can, record the place where it was found.

6

ROCKS OF THE EARTH (2)

SEDIMENTARY AND METAMORPHIC ROCKS

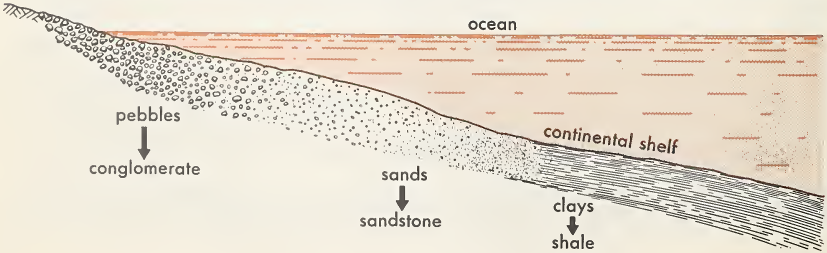
ABOUT 75 per cent of the ***exposed*** surface rocks of the crust are sedimentaries. Nevertheless, they constitute only a relatively thin covering over the underlying igneous rocks. As the name suggests, sedimentary rocks were originally laid down as sediments, usually under water. Unlike igneous rocks, the sedimentaries show bedding or stratification. They are particularly important to the geologist because they give much more information than igneous rocks about earth history. (See Chapter 10.)

Sediments from pre-existing rocks were carried by running water to the continental shelves and deposited there to form sedimentary rocks. In this way, shale was formed from mud and clay, and sandstone was made from sand. The three most plentiful sedimentary rocks are shale, sandstone and limestone. Shale is more abundant than sandstone, and sandstone is more abundant than limestone. These three account for almost 99 per cent of all the sedimentaries.

It is convenient to classify sedimentary rocks into three groups (fragmental, chemical, and organic) according to the origin of the sediments.

Fragmental Sedimentary Rocks. Shale, sandstone and conglomerate are ***fragmental*** rocks; their fragments or particles have been derived from other rocks by erosion. If these fragments are composed of fine particles like clay, the pressure of the overlying deposits is enough to

Fig. 6-1. Sedimentary deposits. Pebbles are compressed to conglomerate, sand to sandstone, clay to shale.



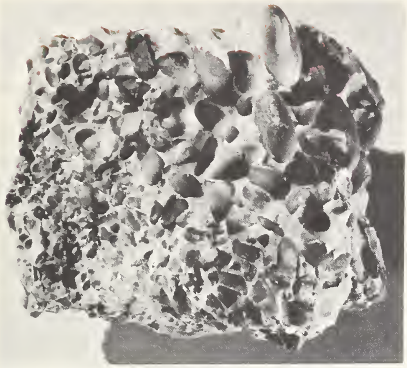


Fig. 6-2. Conglomerate. Notice the cemented pebbles.

make the particles in the lower layer stick together to form a rock. Shale is formed in this way. But if the particles are coarse like sand, pressure from above is not enough to make them stick to each other and a ***cement*** is necessary. The most common cements are silica and calcite. These substances are in solution in sea water, and, under the right conditions, they are deposited on the sand grains and bind them together.

Shale. Shale is made from clay which, in turn, is derived from silicate rocks. In addition to clay, fine particles of quartz, feldspar and even calcite may be present. Since shale is formed from ***fine*** particles, the spaces between the particles after compression are very small. Consequently it is difficult for water to seep through shale, which is therefore a non-porous or ***impervious*** rock. As already stated, shale lacks cement and, as a result, it is a soft rock and is easily weathered.

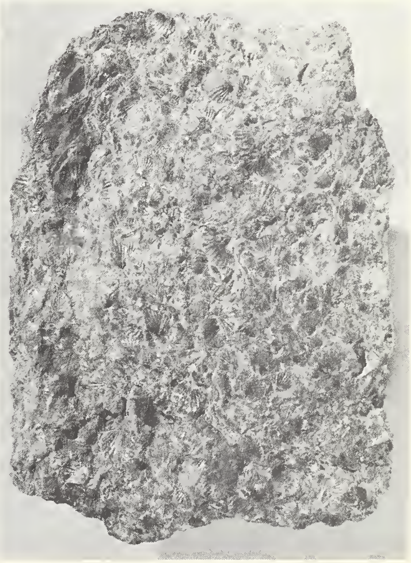
Sandstone. The small grains of sand in sandstone are usually particles of quartz, although grains of feldspar or other minerals are sometimes present. The grains are cemented together, but the cement does not fill the spaces between them. For this reason sandstone is a porous or ***pervious*** rock. It is this property of sandstone that makes it an important underground reservoir for oil. Sandstone may be red, gray, or white in color depending upon the nature of the cement. The color of red standstone, for instance, is due to cement made of iron oxide.

Conglomerate. Conglomerate contains rounded fragments usually quartz, of large size. Pebbles and sand are cemented together so that conglomerate resembles concrete.

Chemical Sedimentary Rocks. Vast amounts of substances in solution are carried to the sea by running water. Under the right conditions some of these substances may be precipitated by chemical action and so form sedimentary rock. For example, calcium salts from plagioclase feldspar react with water and carbon dioxide to form the almost insoluble mineral calcite. Limestones, for the most part calcite, are formed in this way. Most limestones, however, are not pure calcite but contain small amounts of clay and quartz.

Organic Sedimentary Rocks. Organic deposits are composed of the remains of ani-

Fig. 6-3. Shelly limestone. Notice the numerous fossils (brachiopods) embedded in the rock.



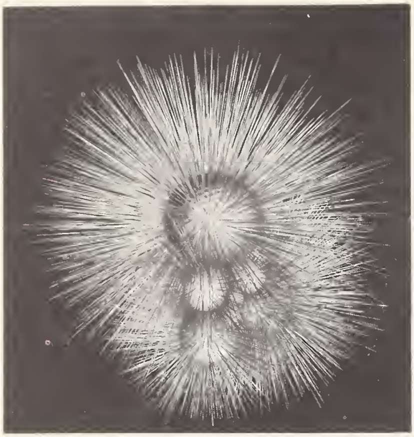
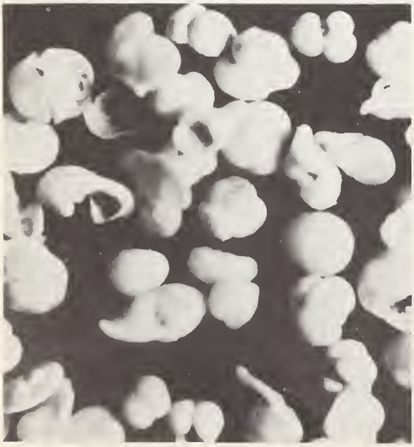


Fig. 6-4. Globigerina builds a many-chambered shell of limestone. This is a photograph of a glass model.

mals and plants. The most familiar example is, of course, coal, a sedimentary rock of vegetable origin important as fuel today. Shelly limestone is still another example. In the deeper waters of the continental shelf shellfish abound in great variety. When the animals die their soft parts decompose and their calcite shells accumulate on the sea floor and are compacted into limestone. If formed beyond the mud zone of the shelf, a limestone consists entirely of fossils. It is therefore evident that limestone may be either a chemical sedimentary rock or an organic rock.

Fig. 6-5. Globigerina shells magnified 25 times.



Coral reefs, which grow in shallow tropical waters, are another form of shelly limestone. Coral deposits are often of great thickness. Indeed, whole islands such as Bermuda in the Atlantic and Wake in the Pacific are formed of coral.

Still another example of an organic sedimentary is chalk, a soft white rock similar in composition to limestone but composed, for the most part of the shells of minute animals called ***globigerina.*** Fig. 6-4 shows a glass model of globigerina greatly enlarged. As you can see, a globigerina skeleton consists of several spherical shells. Notice the long spines through which the living animal takes in food. When the animal dies the spines break off, as shown in Fig. 6-5, and only the spine impressions remain on the skeleton.

Globigerina thrive best in deep water. As a result, chalk is usually free from the particles of mud and sand so often found in limestone. Although globigerina are microscopic in size, the accumulated deposit may be thick enough to form mountains. The white cliffs of Dover, England, are a familiar example of a chalk deposit.

Diatomite. Vast numbers of microscopic plants, called ***diatoms,*** float on the surface of the sea. Covering wide areas, they are the "grasslands" of the sea, and constitute the basic food for fish. Diatoms have skeletons of silica which the plants extract from sea water. If viewed under the microscope these silica skeletons are things of exquisite symmetrical design. (See Fig. 6-6.) Diatoms have flourished through the ages and, in some regions, vast deposits have accumulated and been compacted into a sedimentary rock called diatomite, which is mined in California, Arizona, New Mexico and Nevada.

Diatomite has an unusual structure which accounts for its peculiar properties. It is estimated that one cubic inch of diatomite contains at least 50 million dia-

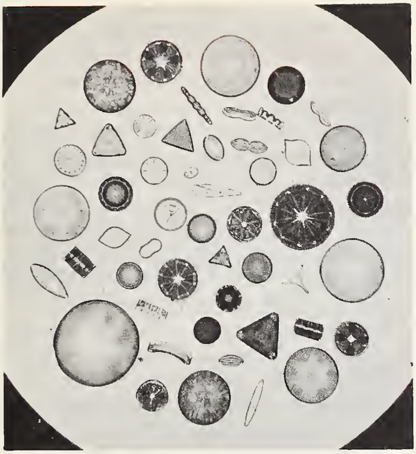


Fig. 6-6. Diatoms are microscopic plants of exquisite design. These are magnified about 100 times.

toms. The spiny, irregular, pitted surfaces prevent the skeletons from packing close together and, in consequence, diatomite is very porous and light. Its specific gravity is only about one tenth the specific gravity of sand. Nobel, of Nobel prize fame, was one of the first to make use of the porous property of this rock. In 1870 he made dynamite by absorbing the highly explosive liquid nitroglycerine in diatomite.

Today diatomite is used for many different purposes. In powdered form it is an effective filter of contaminated water. Even if water is polluted by typhoid and dysentery germs, it is safe to drink after it has passed through a bed of diatomite. Filters of this kind were supplied to troops in World War II. Another unusual application of this powder is its addition to paint used for highway centerlines. The irregular shapes of the diatoms "fluffs up" the surface of the paint and makes it more visible to the motorist.

Metamorphic Rocks. Igneous and sedimentary rocks can be changed by high temperature or high pressure, or both. The resulting rock is called a ***metamorphic*** rock (the word metamorphic means ***changed in form***). Probably the most familiar metamorphic rock is marble, made from limestone or chalk. Other common metamorphic rocks are quartzite, slate, and gneiss (pronounced like ***nice***).

Fig. 6-7. This United States landmark is the Lincoln Memorial, Washington, D.C. The superstructure is of white marble, approximately 80 feet high. It rests on a base composed of three immense marble steps.

Marble. Suppose a mass of molten lava is injected into layers of shelly limestone deep down in the earth's interior. What will happen to the limestone? Some will probably melt, in which case the shell structure and the stratification will disappear. If the melted limestone later cools and crystallizes, it will have the massive appearance of an igneous rock. Rock formed in this way is called marble. It has the same chemical composition as limestone although it is crystalline and much harder than limestone. Marble takes a high polish and is a beautiful building stone. The Lincoln Memorial in Washington, D. C., is a magnificent edifice of white marble. The Statue of Lincoln inside the building is also sculptured out of white marble.

Quartzite. Quartzite is made from sandstone by heat and pressure. It consists entirely of quartz and is exceedingly hard. Unlike sandstone, however, it is not porous. It is one of the most durable rocks and makes satisfactory foundations for large structures.

Slate. If shale is subjected to great pressure it is compressed and hardened. In the process, the mineral particles turn their flat faces in the direction of pressure so that the metamorphosed rock can be cleaved. This is how slate is formed. Slate, as you know, readily splits into thin slabs which are used as roofing material.

Gneiss. Gneiss is a changed granite. It contains the same minerals as granite but, during the change, the crystals were crushed and rearranged in parallel bands. This banded structure is characteristic of gneiss.

Fig. 6-8. Examination of rocks and minerals. The observation of crystals and rock structure is improved by using a magnifying glass.

Demonstration

6-1. Examination of common rocks

Examine such rocks as granite, basalt, limestone, marble, shale, slate, sandstone, quartzite and gneiss. (If possible, use a lens or simple magnifying glass.)

First compare granite, basalt and gneiss. Notice that granite is coarse-grained and basalt fine-grained. Identify the three different minerals in granite. Note the small crystals and the banded structure of gneiss.

Now compare limestone and marble. The limestone may be a grayish calcite deposit, or it may consist of minute shell particles. Marble shows quite clearly the small crystals of calcite.

Now compare sandstone and quartzite. Notice the minute grains of silica in sandstone. It is a porous rock. In quartzite the separate silica grains cannot be seen by the eye.

Finally compare shale and slate. Shale may include fairly coarse particles in the fine background; it may also show bedding. Notice the fine structure of slate and, particularly, its planes of cleavage.

Things to Remember

Sedimentary rocks were originally laid down as sediments.

The ***most plentiful*** sedimentaries are shale, sandstone, and limestone.

Fragmental sedimentary rocks (e.g. shale) are derived from other rocks by erosion.

Chemical sedimentary rocks (e.g. limestone) are precipitated by chemical action.

Organic sedimentary rocks (e.g. coal) are the remains of plants or animals.

Metamorphic rocks (e.g. marble and slate) have been formed from other rocks by heat or pressure.

Questions

GROUP A

1. Name the three classes of sedimentary rocks and give an example of each.
2. What are (*a*) pervious rocks, (*b*) impervious rocks? Give an example of each.
3. What is a metamorphic rock? Give two examples.
4. Discuss some of the industrial uses of diatomite.
5. State whether each of the following rocks is igneous, sedimentary or metamorphic:
   (*a*) granite, (*b*) slate, (*c*) sandstone, (*d*) shale,
   (*e*) obsidian, (*f*) conglomerate, (*g*) gneiss.

GROUP B

6. What is a cement? Why is a cement present in some rocks and not in others?
7. Explain the difference in the formation and composition of chemical limestone and organic limestone.
8. Compare shale and sandstone as sedimentary rocks.
9. Account for the formation of chalk deposits.
10. Explain how limestone may be changed to marble.

Things to Do

Rock collection. Sedimentaries and metamorphic rocks · Make inquiries from a local geologist as to where you can find outcrops of sedimentary and metamorphic rocks. Shelly limestone, for instance, can often be found in a limestone quarry. Your rock specimens will mean much more to you if you yourself collect them in the field. Rock samples should be washed by dipping them in water and scrubbing them with a hand brush.

7

THE CHANGING EARTH

Nothing looks more permanent than a range of mountains; the ancient writer of Genesis called them "everlasting hills." But are hills everlasting? Are the mountain peaks precisely the same as they were a century ago? A century is, geologically speaking, a very short period of time. And so is a thousand years; and so, for that matter, is a million years. Actually there is evidence to prove that the contours of mountains have changed considerably in a million years. Every exposed mountain peak is attacked and slowly worn away. How do we know this? The erosion of mountains and of all land masses is caused by forces of nature, and the most effective of these is the weather. It may seem strange that the sun, the rain, and the wind can, through the ages, cause the destruction of land masses. Yet, as we shall see later, there is no doubt whatever that whole mountain ranges have been completely worn away in this manner.

When geologists speak of this wearing-away process they use the terms ***weathering*** and ***erosion. By weathering is meant the breaking up of bedrock into fragments. Erosion,*** on the other hand, ***implies the transport or carrying away of these weathered fragments.*** In other words, rock material is weathered before it can be eroded.

WEATHERING OF ROCKS

Have you ever noticed that the inscriptions on old buildings, or monuments, or tombstones become almost illegible after one or two centuries? If the carved lettering on a tombstone, for example, dates back to 1860 there are obvious signs of wear; if it dates back to 1760, it may be difficult to read. In the course of 200 years the rock slab may be worn away to a depth of almost half an inch. What would be the effect on the rock slab if the weathering process continued for a thousand years? Or for a million years or more?

Similar visible effects of weathering can be seen on all exposed rocks, particularly in mountainous regions. What causes weathering? The wearing away of rocks is the result of many factors but only three need be considered at this time, namely the effects of temperature changes, freezing water, and solubility.

Weathering Caused by Changes in Temperature. Rock surfaces expand when exposed to the sun's rays. But rock is a poor

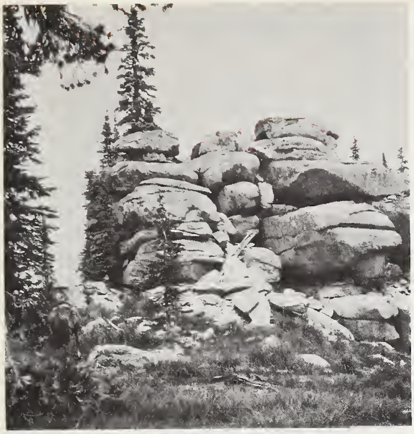


Fig. 7-1. Weathered granite. Temperature changes and ice formation in crevices cracks the rocks; rain rounds them.

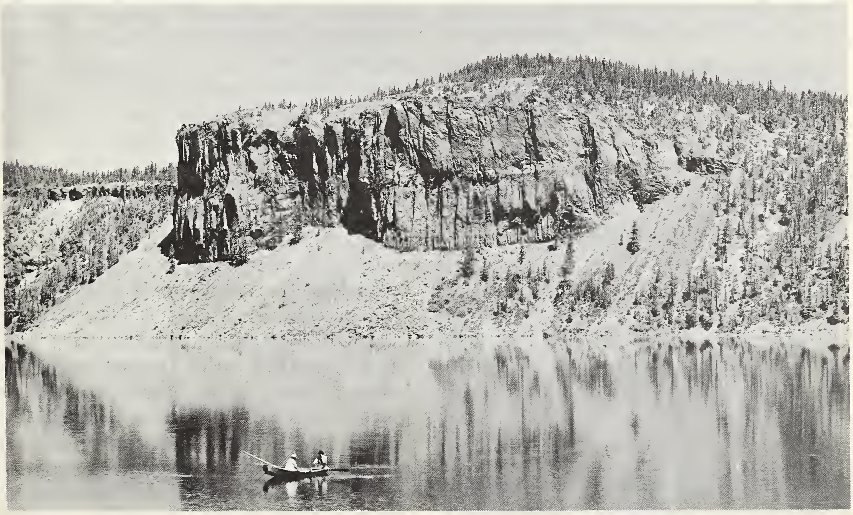
conductor of heat so that the under layers receive little of the sun's heat and do not expand. The expansion of the surface layers sets up internal strains, a process repeated day after day until eventually the surface breaks off. A new surface is then exposed to the sun's heat, expansion occurs, and the flaking-off continues indefinitely.

Or consider the destructive effect of freezing water, in cold or temperate climates. Water seeps into rock crevices or into porous rocks. If the temperature falls sufficiently, the water freezes. From our experience with frozen water pipes we know that water expands when it changes to ice. Actually it expands by almost 10 per cent of its volume. The force of expansion exerts great pressure upon the rock so that the repeated processes of thawing and freezing are enough to split rocks apart and to tear pieces away from the rock mass.

Rock fragments released in these processes slide to the base of a cliff and collect there forming a pile of debris called ***talus.*** Talus can usually be seen at the foot of any cliff in a mountainous region.

Weathering Due to Solubility. Almost all rocks contain some minerals that are somewhat soluble in water. As a result, they begin to crumble when washed by rain through long periods of time. Rain water becomes acidic as it flows over the

Fig. 7-2. Talus is a pile of debris at the base of a cliff.



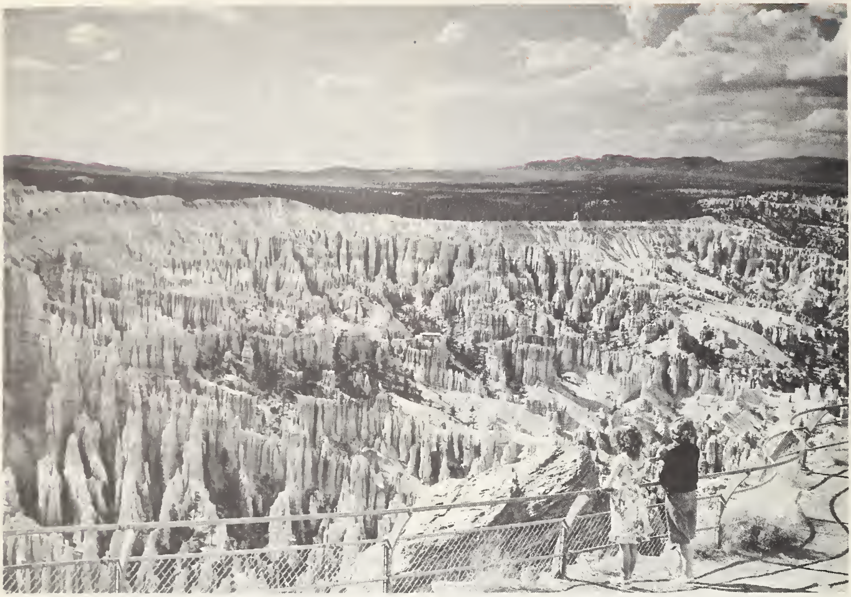
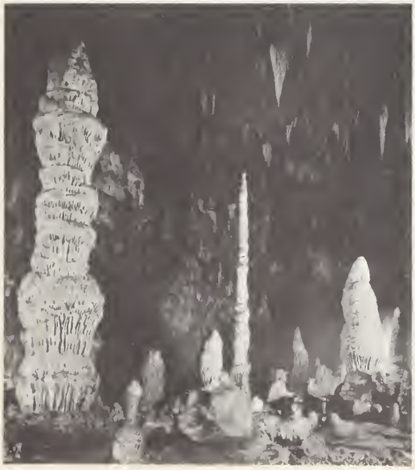


Fig. 7-3. Rain has eroded the limestone rocks of Bryce Canyon, Utah, into fantastic shapes.

earth. You will recall that acids affect rocks, particularly limestone and marble (see page 47). How does rain water become acidic? During the decomposition of vegetation both acids and carbon dioxide are formed. Carbon dioxide dissolves in water to form carbonic acid (see page 177). The acids formed from rotting vegetation are weak and therefore dissolve rocks only very slowly. But over a long period of time the total effect upon exposed rocks is destructive.

Fig. 7-4. Enormous stalagmites in the Carlsbad Caverns. Compare the size of the men in the foreground with the height of the stalagmites.



Although limestone is a non-porous rock, it is usually crisscrossed by cracks and joints. Surface water charged with carbon dioxide slowly percolates through the cracks and along the beds. As a result, limestone rock is dissolved; the cracks become enlarged so that, eventually, a cave or cavern may be formed. You may have seen caverns in limestone regions. The Carlsbad Caverns in New Mexico and the Mammoth Cave in Kentucky are well-known examples of enormous subterranean caves formed by percolating water.

Probably the most impressive features of limestone caverns are the "icicles" of stone called ***stalactites*** and ***stalagmites.*** These are also a solubility effect. As the

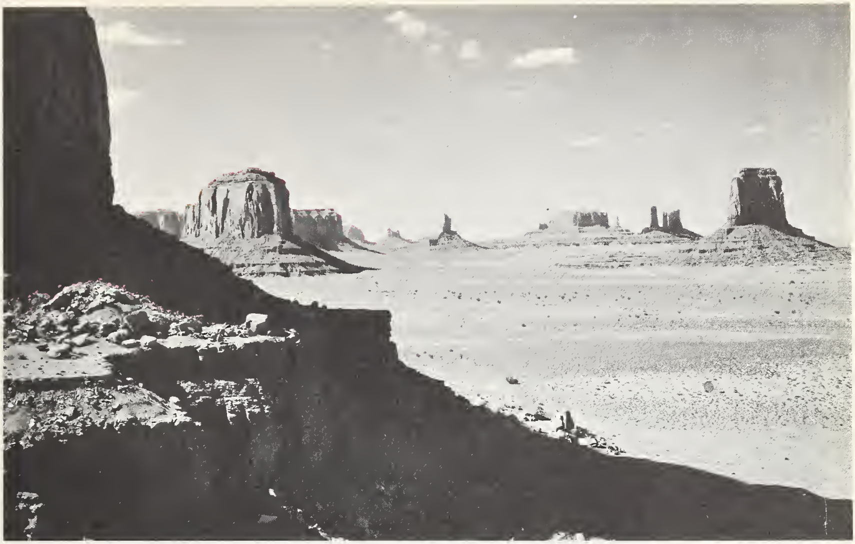


Fig. 7-5. This picture of Monument Valley, southern Utah, shows the effect of wind erosion. The sedimentaries on top of the igneous buttes are almost completely eroded. Eventually the landscape will be a level plain.

water seeps through the limestone it dissolves some of the calcite (calcium carbonate). When the water reaches a cavern it forms droplets that cling to the ceiling. As the water evaporates, the carbon dioxide escapes, and some of the calcium carbonate is deposited as a solid on the ceiling. If this process continues over a long period, the deposit forms a stony "icicle" or stalactite.

Can you now explain how a stalagmite is formed from the floor of a cavern? Some of the water falls from the ceiling to the floor of the cavern. There it evaporates, calcium carbonate is deposited and a stalagmite is started. Sometimes stalagmites and stalactites grow towards each other and eventually join to form a column stretching from floor to ceiling.

Erosion. When we consider the transport of weathered rock material the first thing that probably comes to mind is wind. Winds may pick up dust and sand and carry them great distances, often for hundreds of miles. Wind-blown sand cuts and wears rock surfaces in much the same way that a sandblast etches glass or other hard surfaces. Wind erosion is most effective in dry regions, because there the surface rocks are often bare of vegetation and unprotected. A desert region, for example, is largely the result of erosion by wind. Here the exposed rocks are worn by driven sand, until eventually they are destroyed and a nearly level plain is formed. Wind, however, is by no means the only eroding agent. Running water is far more effective than wind. Indeed, ***running water causes far more destruction of land than all other eroding agents combined.***

Erosion by Running Water. The source of all running water is, of course, rain. The annual rainfall varies a good deal from place to place. In the Nevada desert, for example, it is about 4 inches. Across Canada the annual precipitation varies from over 262 inches at Henderson Lake on the west side of Vancouver Island to 11 inches at Manyberries in southern Alberta. Other annual precipitations are

20 inches in Winnipeg, 42 inches in Montreal, and 54 inches in Halifax. In the far north the precipitation is very low: for example, at Eureka, Ellesmere Island, N.W.T. (80°N latitude) the annual precipitation is only 2.6 inches. The average over Canada and even over the whole earth is about 30 inches per year. Some of this water evaporates, some sinks into the ground, and some runs down surface slopes and returns to the sea. It has been estimated that about 9000 cubic miles (about 10,000,000,000,000,000 gallons) of rain water run back to the oceans every year.

What is the effect of rain and running water on the land? Rain soaks into the ground and collects in the spaces between the soil particles. The water dissolves the soluble constituents of minerals and, as a result, rocks become more porous, their structure weakens and they crumble: sandstone crumbles to sand, and conglomerate crumbles to gravel.

With heavy rains the subsoil becomes saturated, the water begins to flow and, eventually, emerges as a streamlet. On the surface, the streamlet is joined by other streamlets and grows in size. Eventually, it is big enough to move fine particles of soil and sand; lighter materials such as mud and silt are suspended in the water, but heavier materials such as sand and pebbles are rolled along the bed of the stream or river. These larger particles act like sandpaper, deepening the bed and wearing the banks.

The highest point of a river is of course its source, and the lowest point its mouth where it enters the sea. Between these two points tributaries enter the river, and the area drained by the tributaries is called the ***drainage basin***. The drainage basin of the Mississippi, for example, is a vast area of land which extends from the Rockies to the Appalachians. It is important to notice that erosion by running water has two different effects: (1) the river bed is widened and deepened, and, (2) the river basin is lowered. Indeed, in this way rain lowers the land masses of whole continents.

Geologists have calculated that every year the rivers in the United States carry almost a billion tons of rock material to the sea. Some of this is in solution, but most of it is insoluble sediment carried in suspension. The loss of this material from the land is steadily lowering the entire surface of the country at the average rate of about one foot every 8000 years.

Young rivers carve V-shaped valleys in the land. Rain continues to erode the banks so that, in old age, the cross section of a river changes from a steep-sided V to a gently sloping ⌣. If the river flows through dry country there is little rain to widen the valley and, as a result, steep-sided canyons and gorges are formed. The Mississippi and Colorado rivers illustrate these characteristic features of old and young rivers.

The Mississippi. The Central midwestern plain from Ohio to Colorado is drained by the Mississippi River and its tributaries. Through millions of years these rivers have eroded the entire drainage basin, so that it is now almost a level plain. The average lowering of these midwestern plains is about one foot per 5000 years, or about 200 feet in the past million years. The muddiness of the Mississippi is proverbial: Mark Twain is supposed to have said of the river, "It is too thick to navigate but too thin to cultivate." The sediment carried by the Mississippi amounts to more than 500,000,000 tons a year. This enormous load, taken mostly from the fertile lands of Ohio and Iowa, is carried to the Gulf of Mexico. The sheltered waters of the Gulf are unable to carry this sediment very far, so that it accumulates near the mouth of the river as a fan-shaped deposit called a delta (so-named after the Greek letter delta, Δ). The Mississippi delta, which has already extended into the Gulf of Mexico for about 100 miles, is the largest in the world.

As one might expect of such an old river, its bed has almost reached base level

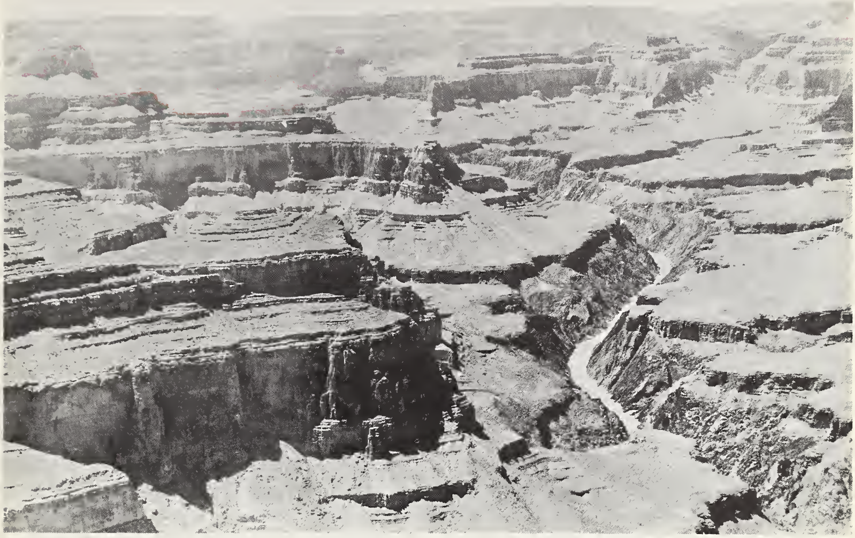


Fig. 7-6. The Grand Canyon. Notice the horizontal strata in the wide canyon, the talus at the base of cliffs, and the Colorado River at the bottom of a deep gorge.

and its valley walls are shallow. Consequently, the river frequently overflows its banks and disastrous floods sometimes occur, particularly during spring rains. When the flood waters meet the land, they deposit much of their load (why?), forming broad ridges of earth, called levees, parallel to the stream. To keep the river in its channel during the spring floods, engineers have strengthened and enlarged these natural levees.

The Colorado River. Another example of river erosion is the valley of the Colorado River which has its source in the Rockies and flows through Arizona, Nevada and southern California. The Colorado is a much younger and swifter river than the Mississippi, and during the relatively short period of a million years it has carved a canyon a mile deep out of solid rock. Only unusual geological conditions made it possible for the river to cut its bed through this great depth. It so happens that throughout most of the period of erosion the land mass was slowly rising, which, in effect, resulted in the lowering of the base level of the river. Imagine a layer cake being cut by raising the cake through the knife, rather than by lowering the knife through the cake. This is the kind of cutting that enabled the Colorado to carve several deep canyons in its rocky bed.

The Grand Canyon in northern Arizona, the largest of the Colorado canyons, is a gigantic trough almost a mile deep and about 200 miles long. Its width varies from 4 to 18 miles. The rim of the canyon is about 6000 feet above sea level. The rocks of the cliff wall are sedimentaries resting upon igneous rocks. In making the canyon, the Colorado River first cut through about 4000 feet of sedimentary rocks consisting of red sandstone, white sandstone, and gray-blue shale. Then it carved a gorge about 1000 feet deep in

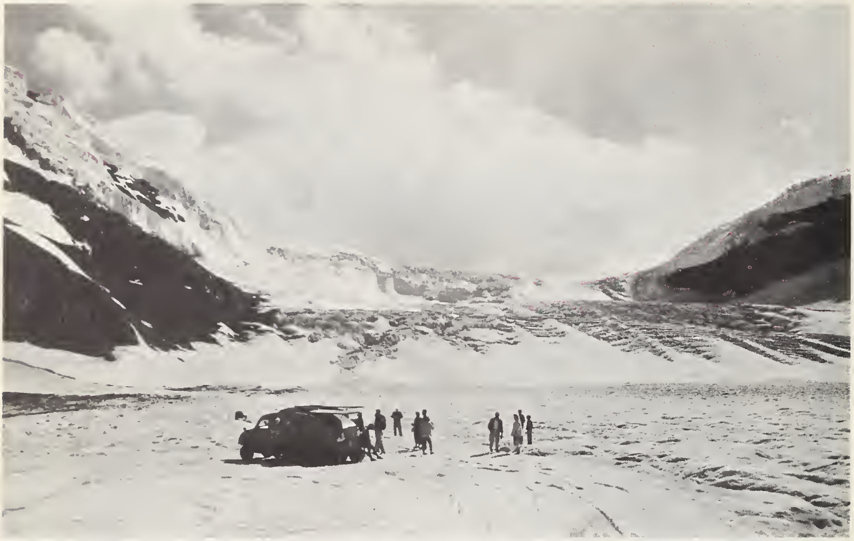


Fig. 7-7. The Athabaska Glacier, an outlet of the Columbia Ice Fields, showing an icefall in the background. The people are standing on snow-covered ice at the foot of the glacier, just above its terminus.

the igneous rocks below the shale. Every year thousands of visitors are impressed and even awed by this spectacular demonstration of river erosion.

Erosion by Ice—Valley Glaciers. Some regions of the earth are above the snow line. Here the snow that falls during the winter does not completely melt during the following summer, so that these regions are covered by perpetual snow. Does snow pile up indefinitely above the snow line? As snow accumulates its pressure increases and the lower layers of snow are compressed to ice. The ice begins to flow under its own weight and it is squeezed through the valleys to a lower level. A river of ice formed in this way is called a ***valley glacier.***

Although limited in scope, glaciers are powerful eroding agents. As a glacier slowly moves down a mountain it gouges rock fragments out of the sides. These fragments may fall through cracks in the glacier, become embedded in the ice, and are crushed to boulders, pebbles, sand, and silt. Thus the bottom and sides of a glacier are armed with "teeth" which scour the rock as the glacier moves along. Sand and silt smooth and polish rock surfaces; coarser rock fragments cause scratches, or striations as the geologist calls them. Both these effects are characteristic of glacial action. As the valley is gouged wider and deeper the walls are undercut. As a result, enormous boulders as well as much loose material collect on the surface and are carried along.

The valleys carved by glaciers are straight, wide, and U-shaped, quite unlike the valleys eroded by rivers. These characteristics, together with the smooth, striated rock surfaces, make it easy to recognize the work of glaciers.

Where in its downward path does a glacier melt? It eventually reaches a point where the ice melts as fast as the glacier advances. This is its lower limit and is well below the snow line. At this point all the rocky material carried by the glacier is deposited as a terminal ***moraine.*** If the glacier melts in the same region for a long time, the deposits may become

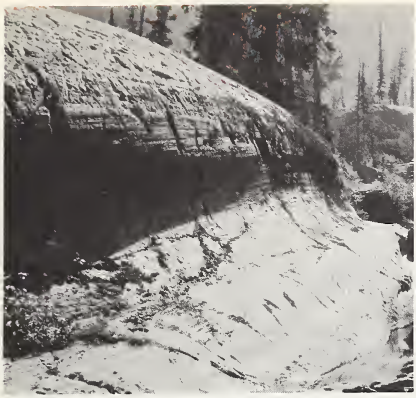


Fig. 7-8. The rock wall of Glacier Creek, Flathead National Forest, Montana. Notice that the wall shows characteristic glacier action; it is smoothed and polished, scratched and grooved.

hundreds of feet thick. Glacial deposits formed in this way are found throughout southern Saskatchewan and central Alberta. They can easily be recognized because the boulders, pebbles and clay are all mixed together; they do not form layers as in sedimentary rocks.

Valley glaciers are found mostly in the Himalayas, the Andes, the Alps, and Alaska. In Canada, valley glaciers are common in the Rocky Mountains. The best known of these valley glaciers form part of the Columbia Ice Fields. Farther north, in the Yukon Territory and Alaska, the snow line is down to 5000 feet. There glaciers are very common, the largest of them being about 40 miles long and 1000 feet thick. Many of them reach the sea, break up, and float away as icebergs.

Continental Glaciers. Suppose the snow line is near sea level. Would valley glaciers then be possible? In such a case large areas of land would be covered by a vast, thick sheet of ice, called a ***continental glacier.*** There are only two continental glaciers, one in the Northern and one in the Southern Hemisphere. In the north, Greenland, except for a small strip of land near the coast, is covered by ice which in places is more than 1000 feet thick. In the south, the continent of Antarctica is almost completely covered by an ice field that extends for millions of square miles, and varies in thickness from 4000 to 9000 feet.

The vast Antarctica ice cap accounts for about 85 per cent of the glacial area of the earth, and the Greenland ice cap about 10 per cent. Thus all valley glaciers make up not more than 5 per cent of the earth's glacial area. It has been estimated that if all the glaciers were to melt, the level of the oceans would rise by 100 to 200 feet. This figure, however, is uncertain because the depth of the Antarctica ice cap has not been thoroughly explored.

There is no doubt that continental glaciers have been much more extensive in past geological ages. The evidence shows that these ice sheets covered the tops of mountains, and the peaks of these submerged mountains were rounded and smoothed by the slowly moving ice.

THE ICE AGES

Throughout most of its long history, the earth has probably been free of glaciers. During the past million years, for some unexplained reason, there have been at least four ice ages; that is, ice sheets have advanced and retreated four times. We are now in the thawing stages of the fourth ice age which began its retreat about 25,000 years ago. During the severe stage of the fourth ice age there were no fewer than five continental ice sheets—in North America, in northern Europe, in Siberia, and also in Greenland and Antarctica. The first three have now disappeared but the other two, of course, still remain. At their maximum advance, ice sheets covered at least 30 per cent of the earth's surface. Since this ice was formed from the waters of the oceans, sea level fell far below its present height.

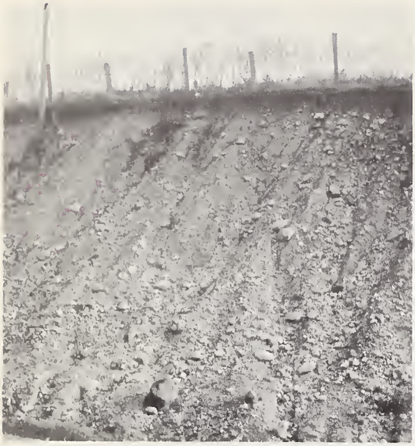


Fig. 7-9. A glacial deposit. Notice that boulders, pebbles, and clay are all mixed together.

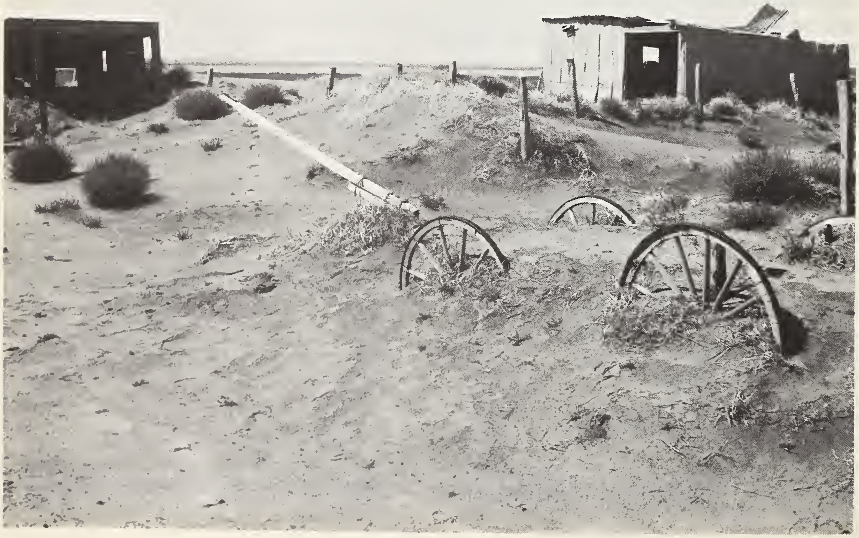
Indeed, geologists tell us that sea level was probably more than 300 feet lower than it is now. If this is true, it would have been possible to walk to the site of the present "floating radar stations," 100 miles or so east of Cape Cod.

Ice, thousands of feet thick, covered Canada, extending into the United States as far south as New York City, St. Louis, and along the Missouri River. When the ice melted, ridges of sand and gravel (moraines) were left. There are many such moraines in Canada: one extends across the countryside just north of Toronto and there are others in southern Saskatchewan and central Alberta.

The effect of the glaciation on Canada was considerable. Before the glaciation such lakes as Winnipeg, Manitoba, and Winnipegosis did not exist. Indeed these lakes owe their beginnings to the glaciation. The glaciation also dammed up many old rivers by filling their valleys with debris, forming the great numbers of lakes now existing in the Canadian Shield. The other large lakes of Canada were greatly modified.

The prairie provinces and the southern portion of Ontario were enriched by the finely ground rocks scraped from the Canadian Shield. The deep coulees that

Fig. 7-10. Destruction of farmland by wind erosion. Soil has collected in the fence rows to a depth of several feet.



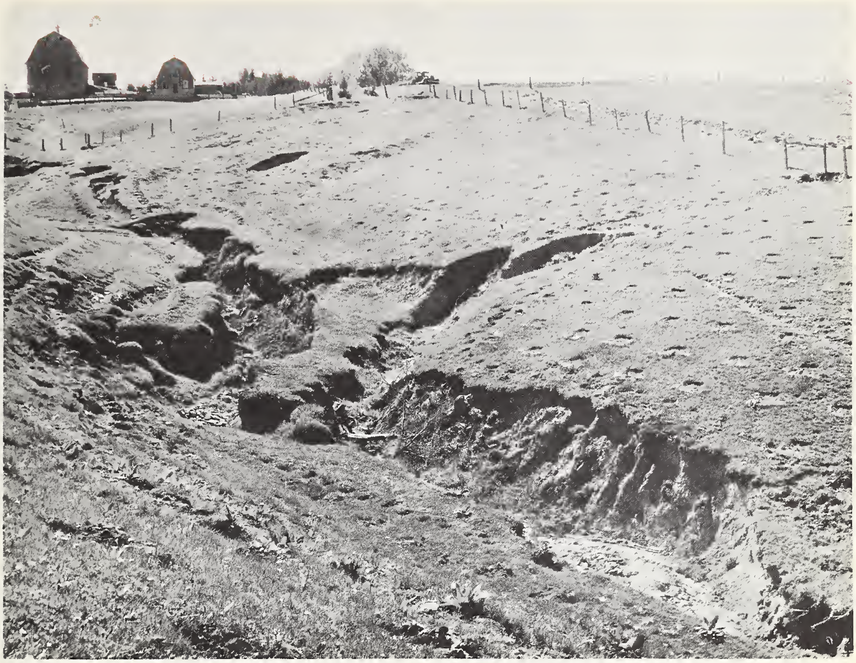


Fig. 7-11. A badly eroded pasture showing the deep gully formed by the rapid run off of surface water after a rainstorm.

cut the prairies were formed by rivers from the melting glaciers.

The mountains of eastern Canada and the northeastern United States were completely covered with ice. Their summits therefore have been smoothed and rounded. There are no knife-edge ridges as in the western mountains.

The western mountains were eroded by many valley glaciers. They were not immersed by the ice sheet. As a result, the peaks of the Rockies are much more rugged than the peaks of the eastern mountains.

Soil. Have you noticed that the natural growth of trees and plants changes from place to place? In one region pine and fir may predominate; in another, maple; in another, oak; in still another, rhododendron. How can you account for these differences? They are due largely to differences in the chemical composition of soils.

The formation of soil is in large measure due to the weathering of rocks. Rocks vary greatly in their resistance to weathering. Sandstone and quartzite are durable rocks; their cements disintegrate, but the quartz itself weathers exceedingly slowly. Shales, on the other hand, are the weakest of rocks and disintegrate most easily. Granite weathers slowly and its minerals, feldspar and mica, change chemically during this process. Feldspar, for example, changes to clay.

If the bedrock is almost horizontal the weathered particles are not easily carried away, but remain to form soil. Soil, however, is not merely disintegrated rock. Microscopic organisms called bacteria bring about chemical changes in the soil that make it possible for plants to live.

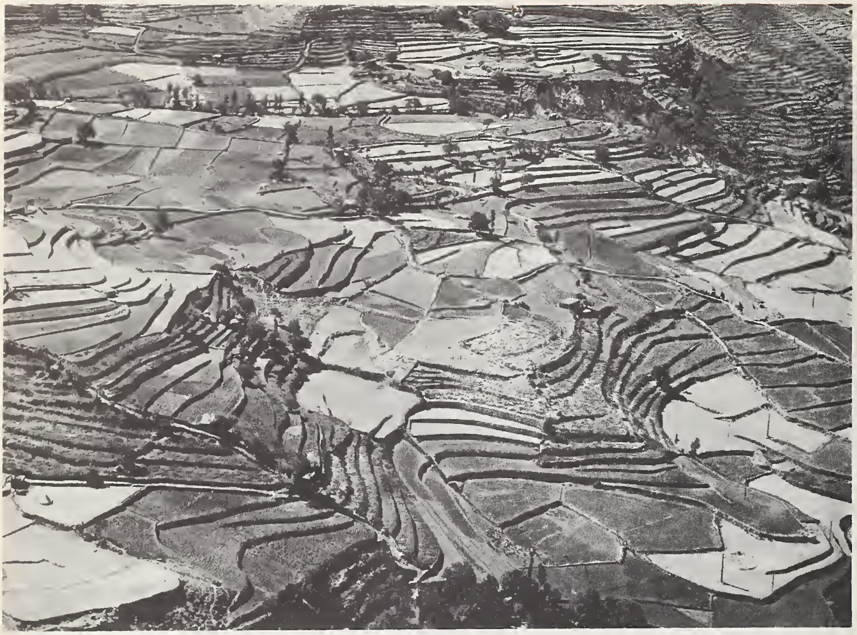


Fig. 7-12. This remarkable photograph shows the terracing of a steep valley in Lebanon, Syria.

Soil is also enriched by the remains of vegetation that has lived and died there. Animals such as earthworms also enrich the soil by boring tunnels, which convey air and water to the soil particles.

The depth of soil varies considerably, depending on the slope of the ground and other factors. In the temperate zones soil usually ranges from two to five feet in thickness, although the rich topsoil is seldom more than a foot deep.

Soil Erosion and Conservation. Soil is a product of geological and organic processes extending through centuries of time. Without soil plants could not live; it is one of the indispensable natural resources of the earth. In the past, vast tracts of valuable topsoil have been carried away by wind and rain, and only within recent times have effective steps been taken to prevent these losses.

Before the West was first settled, the prairies were covered by buffalo grass and other native grasses. Once the grass was removed, the topsoil was exposed to high winds which, in these arid regions, had a destructive effect. Although some native grass has been replanted to halt the erosion, parts of Kansas, Oklahoma, and Texas are still known as the "Dust Bowl." In this region severe dust storms still occur from time to time. One of the worst took place in 1934 when, after a long drought, high winds carried away more than 300 million tons of fertile topsoil, much of it being lost in the Atlantic Ocean.

Water also washes away topsoil, particularly from land with a marked slope. The steeper the slope, the greater is the speed of surface water after a rainstorm, and the greater is the loss of soil. How can such erosion be retarded, and how can wasteful losses be prevented? This problem is now being attacked by the Department of Agriculture and by vari-

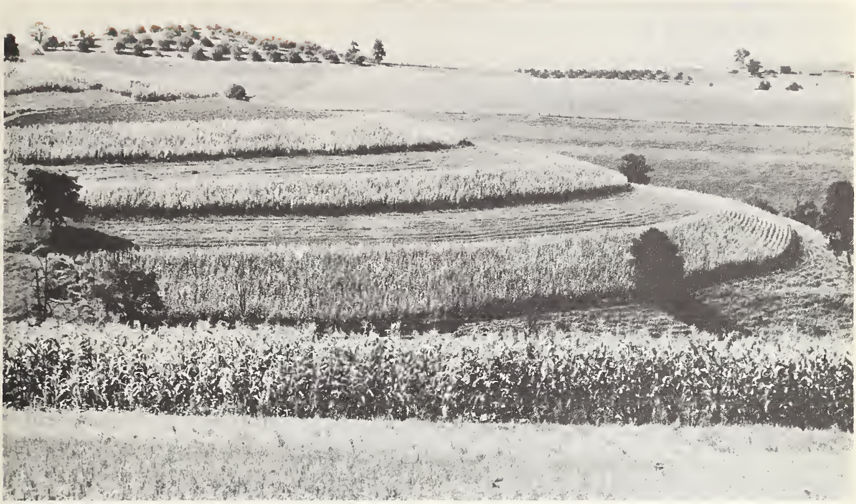


Fig. 7-13. Contour strip cropping on sloping land. Note the alternate beds of wheat and corn along the contour of the slope.

ous agricultural colleges as part of a ***soil conservation*** program throughout the country. In general, soil can be conserved by improving farming and lumbering practices—by the reforestation, or by terracing, contour plowing, and strip cropping on land used for agriculture.

Terracing prevents erosion by breaking up a long slope into a number of shorter and gentler ones. It is the oldest and most effective method of soil conservation; it was used by the Incas in Peru, and the Chinese and Japanese have used it for centuries. In terracing, the slope is changed to a series of steps supported by retaining walls, so that the soil is retained on horizontal ledges and runoff is thereby largely prevented. Retaining walls are expensive to build, however, and parallel earth ridges are now usually preferred.

Contour plowing is the modern adaptation of terracing by earth ridges, and is frequently used by farmers of hilly land. A series of furrows is made, each furrow being at a constant elevation. Every furrow then acts as a reservoir to retain water which can go into growing plants instead of running into streams.

Sometimes rows of crops such as corn, potatoes, and tobacco, are alternated with rows of grain crops such as wheat and oats. The roots of wheat and oats are fibrous and grow close together. They therefore hold water and even supply enough for the crops in the alternating beds. This alternation of different types of plants is called ***strip cropping.***

Things to Remember

Weathering is the disintegration of exposed rock. Weathering may be caused by changes in temperature of surface rocks or their solution in water.

Erosion is the carrying away of weathered rock material.

Running water is the most effective eroding agent.

A ***glacier*** is the most powerful eroding agent.

A ***valley glacier*** is a river of ice.

A ***continental glacier*** is a vast sheet of ice covering the land. North America was covered by a continental glacier, 25,000 years ago. It extended as far south as New York, St. Louis, and the Missouri River.

Terracing, contour plowing, and strip cropping are methods of ***soil conservation.***

Questions

GROUP A

1. Name three methods by which rocks may be weathered.
2. Name three eroding agents.
3. State the annual precipitation in (*a*) the Nevada desert, (*b*) Halifax, (*c*) Montreal.
4. Explain the terms (*a*) delta, (*b*) glacier, (*c*) levee.

GROUP B

5. Discuss temperature as a weathering agent.
6. Explain the formation of caverns in limestone regions.
7. "Erosion by running water widens and deepens a river bed, and also lowers the whole river basin." Explain what this means.
8. What are the main differences between the Mississippi and Colorado as eroding agents?
9. Compare and contrast valley glaciers and continental glaciers.
10. What was the effect of the glaciers of the fourth ice age on the land mass of Canada?
11. Discuss two methods of soil conservation.

Things to Do

Models of land features · Simple models showing not more than one or two features are very instructive. The model should be small (not more than 9 in. x 9 in., and 3 or 4 in. in height), and may be made in modelling clay or plaster of Paris. River valleys, deltas, glaciers and moraines can all be shown in this way.

The best models can be retained in permanent form by making plaster casts of them. This is done by pouring plaster of Paris mixed with water onto the model used as a mold. Remove the cast when it has hardened. The cast must now be used as a mold. To do this, cover it with vaseline and pour more plaster of Paris into it. This final cast is a replica of the original model, and its appearance can be improved by applying suitable paints.

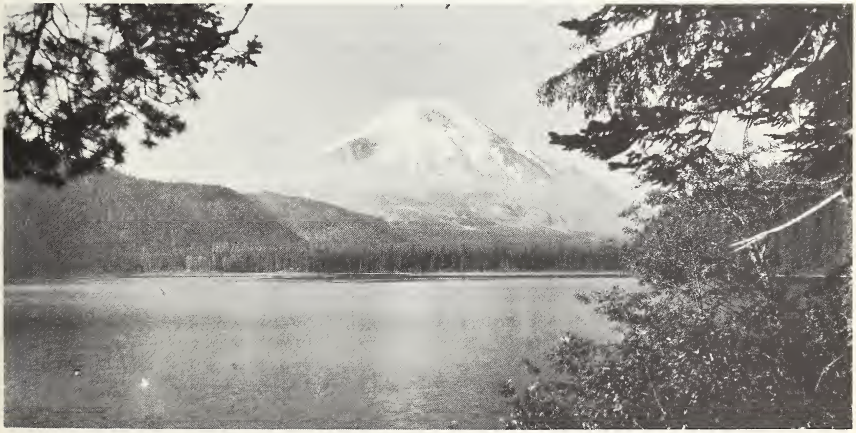
8

THE FORMATION OF MOUNTAINS

WEATHERING and erosion are forces which attack, and ultimately destroy, every exposed rock. Suppose these destructive forces operated without hindrance through the ages, what would be their ultimate effect? All mountains would be levelled, and even the plains would be washed into the sea; there would be no abode on earth for mankind. Fortunately there are forces which lift up the land and so offset the effects of erosion. They are called constructive forces.

These constructive forces may operate very slowly by warping the crust into enormous folds called mountains, or they may produce sudden disastrous effects such as earthquakes or volcanic activity. There is plenty of evidence to show that the tops of mountains were once at much lower elevations. For instance, there are marine fossils in the sedimentary rocks of the Rockies, which proves that the Rockies once formed the floor of an ancient sea. How did these fossils manage

Fig. 8-1. Mount St. Helens in the Cascade Mountains, Washington. Mount St. Helens is a volcanic peak about 14,000 feet above sea level, and well above the snow line.



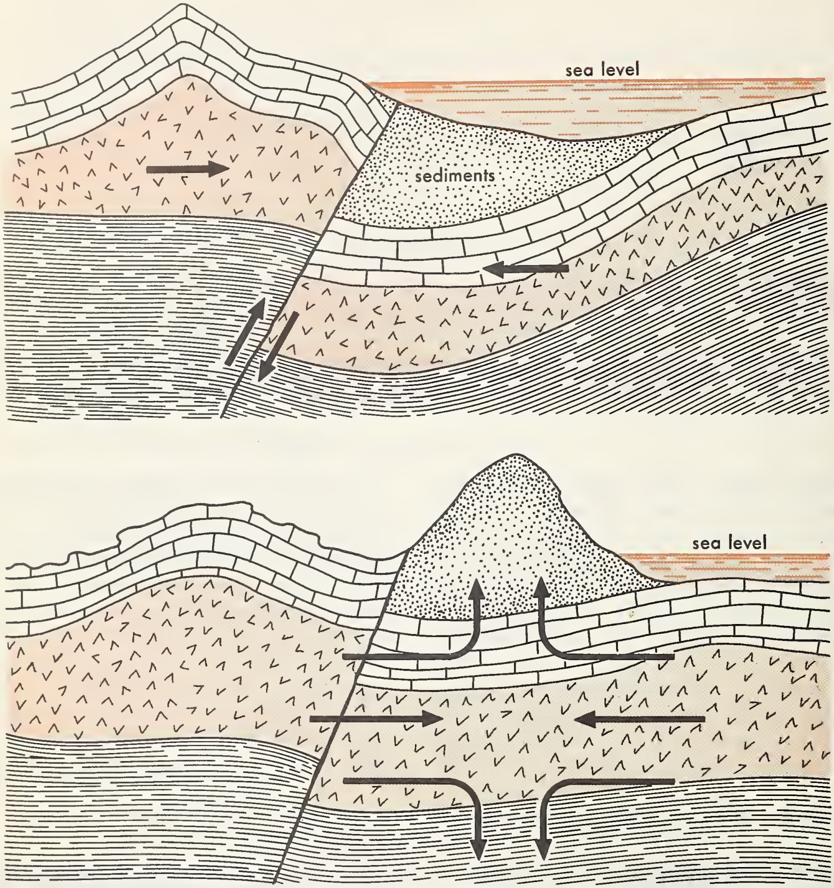


Fig. 8-2. A geosyncline. Pressure of sediments causes slippage along the fault. As a result, sediments of enormous thickness accumulate. Eventually the horizontal stresses force up the sediments as a range of mountains.

to rise thousands of feet above sea level? Geologists tell us that, in the distant past, a shallow sea cut through the continent of North America all the way from Hudson Bay to the Gulf of Mexico. As the adjacent land was eroded, the sediments were carried into this shallow sea; then, in the course of time, these deposits were, in some way, forced high above their original level to form the Rocky Mountains. How can we account for the stupendous forces needed to raise mountains? It is believed that the cooling of the earth's crust and the formation of geosynclines play an active part in this process.

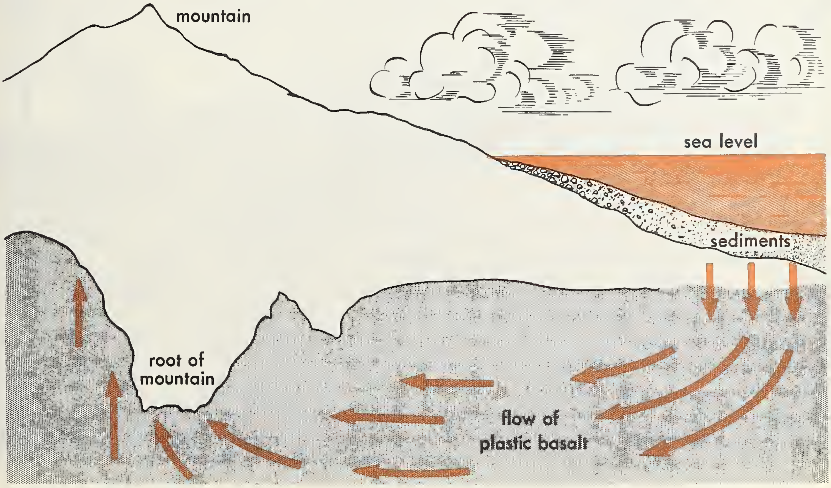
Cooling of the Earth's Crust. The uppermost part of the earth's crust has been a rigid shell, about 30 miles deep, for the past three billion years or so. During this period of time rock to a depth of 200 to 300 miles or so slowly cooled and contracted. Thus the crust became too large for the interior and the rocks of the crust were therefore subjected to enormous stresses. It was stated on page 43 that, even at depths of four or five miles, rocks are very hot. Moreover, they are compressed on all sides by surrounding rocks. The combined effects of heat and pressure make the rocks plastic so that they flow and bend rather than break. Thus the rocks of the crust were slowly bent upwards because, in this way, they occupied less area and so could accommodate themselves to the shrinking interior.

Hundreds of cubic miles of solid rock have been raised by the force of contraction to form plateaus and mountains. However, the cooling of the earth does not provide the complete answer to the formation of mountains. If we examine the positions of mountain ranges such as the Andes and Rockies in America and the Himalayas in Asia, we are led to conclude that most of the great mountain chains of the earth have been formed along the borders of the continents. Why is this? To answer this question we must first make a brief reference to ***geosynclines*** and ***isostasy***.

Geosynclines. Geologists are agreed that all major mountain ranges are formed from geosynclines. What, then, is a geosyncline? It is a trough or basin in shallow water in which sediments from the land collect for millions of years. As the sediments accumulate the trough is continually deepened so that deposits, thousands of feet thick, are formed. Ultimately, the deepening process reaches a limit, and the sediments are forced up as a mountain. These processes can be illustrated by reference to a particular geosyncline that gave rise to the Appalachians, which were once marine deposits six or seven miles thick.

The Appalachian trough was covered by an inland sea which extended from Newfoundland to the Gulf of Mexico. The trough or geosyncline was deepened by compression from the sides, and sediments accumulated from land masses on the east and west. When the sediments and the trough bottom reach great depths (far greater than the depth of the sea), the heat causes them to expand and sometimes melt. Then they are unable

Fig. 8-3. Isostasy. The weight of the sediments causes the underlying rock to flow, thus buoying up the mountain. In this way, a state of balance between the highlands and the lowlands tends to be preserved.



to push further into the earth. As compression continues to be applied from the sides of the geosynclines, the sediments are forced up and, eventually, form a mountain chain. The Appalachians rose from the sea in this manner about 200 million years ago. New mountains are at once attacked by the forces of erosion but are restored, at least in part, by the process of isostasy. (See Fig. 8-2.)

Isostasy. Mountains formed from geosynclines have roots which extend far into the plastic basalt below the mountain. Indeed, a mountain resembles an iceberg: an iceberg floats in water; a mountain floats in the heavier rock. This state of balance between the mountain highlands and the surrounding lowlands is called ***isostasy.*** As the mountain is eroded and the sediments carried to the sea, the balance is disturbed. What happens? If part of the top of an iceberg falls away, the berg is buoyed up by the water and rises so that about the same amount of ice still emerges from the water. The same is true of the mountain. The accumulated sediments cause the underlying plastic basalt to flow and, as a result, the mountain slowly rises to adjust the imbalance. In this way isostasy tends to offset the wearing away by erosion by restoring, at least in part, the elevation of the mountain.

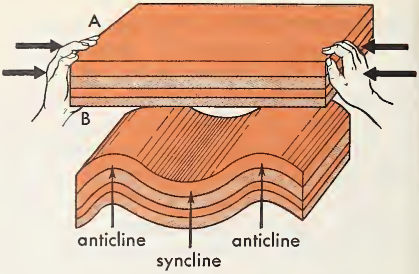
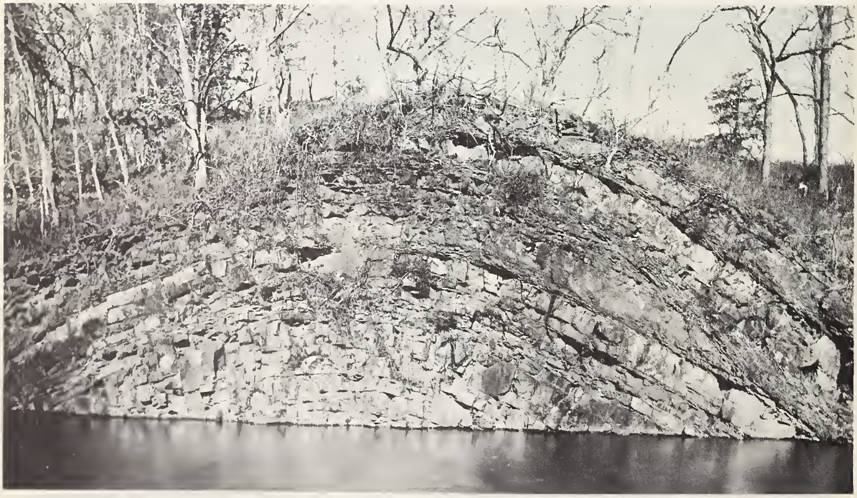


Fig. 8-4. Sedimentary strata may be folded by horizontal pressure into parallel ridges and furrows called anticlines and synclines.

With this introduction to the mountain building process we can now consider the three main types of mountains, namely, ***folded, block,*** and ***volcanic*** mountains.

Folded Mountains. As already stated, folded mountains rose from the floors of ancient seas. Horizontal layers of sedimentary rocks were squeezed in a geosyncline and folded into parallel ridges and

Fig. 8-5. An anticline is an upfold in rock strata.



furrows and ribbed like a washboard. Folds are sometimes so large that a single fold may extend for many miles; or, on the other hand, they are so small that several folds may be observed in the same cliff or quarry. The upfolds are called ***anticlines,*** and the downfolds are ***synclines.***

The largest mountain chains on earth are folded mountains, and marine fossils are usually found on their summits. For some reason, igneous activity always occurs whenever elevation and folding from a geosyncline take place. As a result, the core of almost every mountain structure is a mass of granite. The Appalachians in the east, and the Rockies and the Coast Ranges in the west, are examples of folded mountains. Notice they all run parallel to the coast, and at some time in their history they were all near the coast.

The original Appalachians were uplifted at the end of the Paleozoic era (see page 93), about 200 million years ago. The great folds of the ancestral Appalachians were the anticlines and probably resembled the Alps in elevation and grandeur. During the Mesozoic, a period of about 140 million years, the anticlines were completely eroded down to a plain. This plain was then uplifted to a plateau at the end of the Mesozoic era, about 60 million years ago. This plateau has been dissected by erosion to form the present Appalachians. From this geologic history we are therefore led to the strange but true conclusion that the tops of the present Appalachians are part of the synclinal folds of the original Appalachians.

The sediments that formed the Rockies were also laid down in a sea that covered the interior of North America. The Rockies are much younger than the Appalachians and were raised at the end of the Mesozoic, some 60 million years ago. During the uplift folding took place, and masses of granite constitute the cores. Erosion has exposed the granite cores on many of the high peaks, and sedimentaries rest upon the core at a lower altitude.

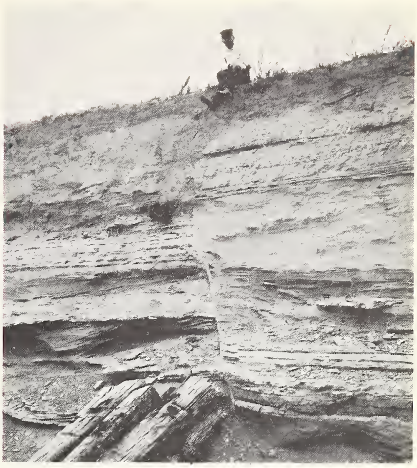
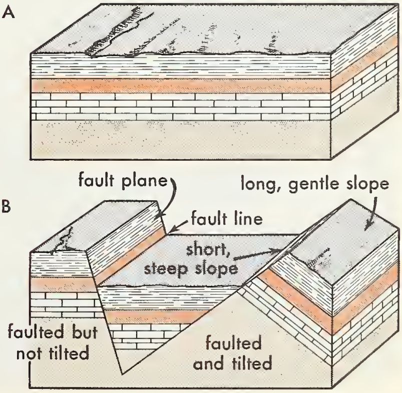


Fig. 8-6. A fault. Rock layers have been displaced downward on the left side of the fault.

Block Mountains. Rocks can be folded in anticlines and synclines by lateral pressures set up by earth movement. Sometimes, however, pressures cause rocks to split, and the rock layers on one side of the break may slide up or down. An arrangement such as this is called a ***fault.*** Notice in Fig. 8-6 that each layer on one

Fig. 8-7. The diagram B shows two faults. The strata on the left are the undisturbed layers; those in the center have slipped along the fault but are still horizontal; those on the right have been faulted and tilted.



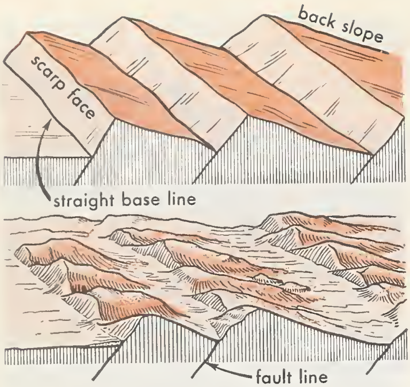


Fig. 8-8. Block mountains may be formed if large areas of crust are broken by faults and tilted to form enormous ridges.

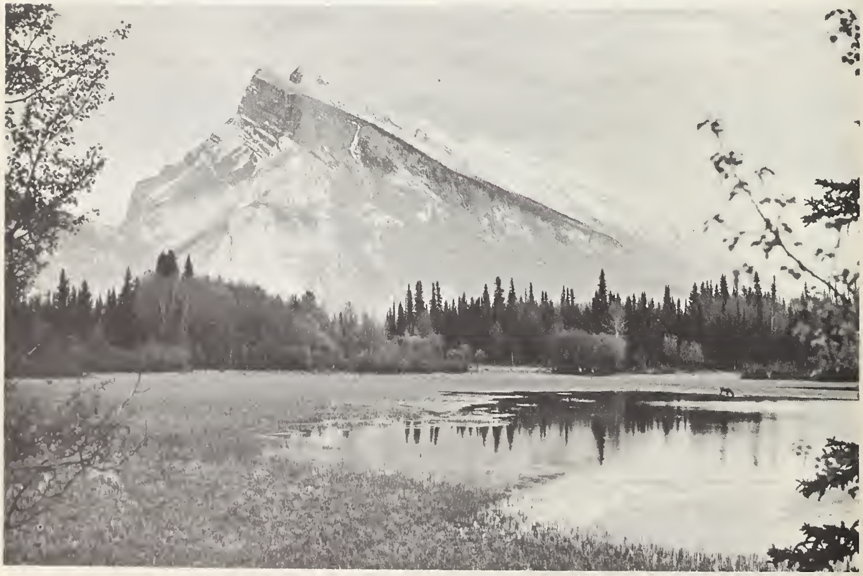
side of the break, or fault plane as it is usually called, is on a different level from the corresponding layer on the other side.

Internal pressures may sometimes cause a vast area to be bowed up into a great anticline. The top of such an anticline is stretched and may become broken by faults hundreds of miles long to form massive blocks as shown in Fig. 8-8. When large blocks of this crust are faulted and tilted, enormous ridges are formed which, when eroded, become mountain chains. These are block mountains. Notice particularly that block mountains have a characteristic feature. They have a short, steep slope on one side and a long, gentle slope on the other.

Faulting on a stupendous scale has occurred in Utah and Nevada. The Nevada faulting has resulted in the Sierras, the largest block mountains in the United States. Actually they are a single block more than 400 miles long and 60 miles wide. Its gentle slope faces west; its eastern wall is a gigantic rock scarp which towers two miles above the desert at its foot. In the Rocky Mountains many of the high peaks are formed by fault blocks. Here, however, they are pushed over each other.

Sometimes an enormous block of the earth's crust is elevated and remains almost horizontal, to form a plateau. The

Fig. 8-9. Mount Rundle in Banff National Park is an example of block fault.



Colorado Plateau, for example, in the southwestern part of the United States, is an elevated plain formed largely by faulting. Its average elevation is about 8000 feet above sea level. As explained on page 63, the Colorado River has cut gorges into this plateau. The rocks exposed in these strata are almost horizontal, proving that almost no folding accompanied the elevation in this region.

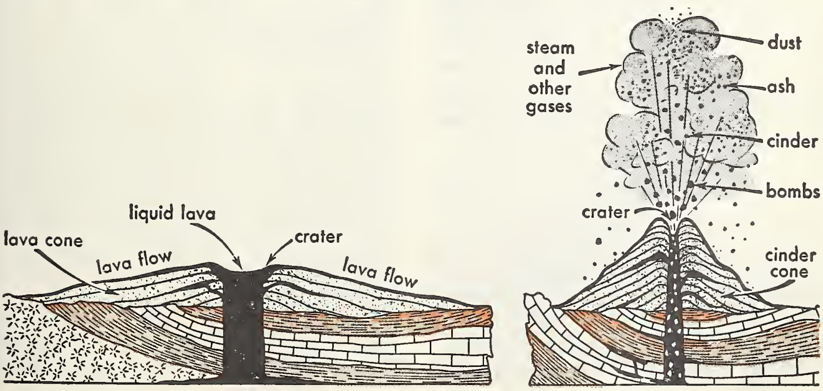
Volcanoes. Suppose a crack extends from the crust into the interior of the earth, say for a distance of 30 miles or so. The enormous pressure exerted upon rocks at this depth forces some of the hot plastic rock up the crack. As the rock rises, it is relieved of the pressure of the overlying layers and, as a result, it melts. The molten rock may escape at the surface as red-hot lava to form a volcano. As the lava and ash pile up about the opening they form a conical heap. The heap becomes higher and higher with successive eruptions, so that eventually a mountain is formed.

Although active volcanoes are rare in Canada and the United States, more than 500 of them are scattered over the earth. In addition, there are several thousand extinct or dormant ones. Active volcanoes are of two types, ***quiet*** and ***explosive.*** They differ markedly in their characteristics. The difference is due largely to the chemical composition of the rock lava. You will recall that granite has a good deal of free silica, whereas basalt has none. This difference in composition affects the melting point (or solidifying temperature) of the rock; the melting point of basalt is considerably lower than that of granite. As a result, basalt lava remains fluid at lower temperatures than granite lava; consequently it flows more easily. Basalt lava is associated with quiet volcanoes.

Quiet Volcanoes. Since molten basalt flows freely, it spreads over a wide area and forms a cone with a gentle slope, before hardening to basalt, obsidian, or pumice (see page 49). (The spongy texture of pumice is due to bubbles of steam, carbon dioxide, and hydrogen sulfide still present in the lava when it hardens.)

The Hawaiian volcanoes, including Mauna Loa, are the best examples of quiet volcanoes. Mauna Loa, which is still active, has taken millions of years to build the largest volcanic cone on earth. It rises 30,000 feet above the sea floor and 14,000 feet above sea level; that is, from base to summit it compares with Mount Everest, which towers 29,000 feet (about six miles) above sea level. The circumference of the base of Mauna Loa

Fig. 8-10. Quiet volcanoes (as shown on the left) form a gentle slope and spread over a wide area. Explosive volcanoes (as on the right) form steep-sided cones of ash and cinders.



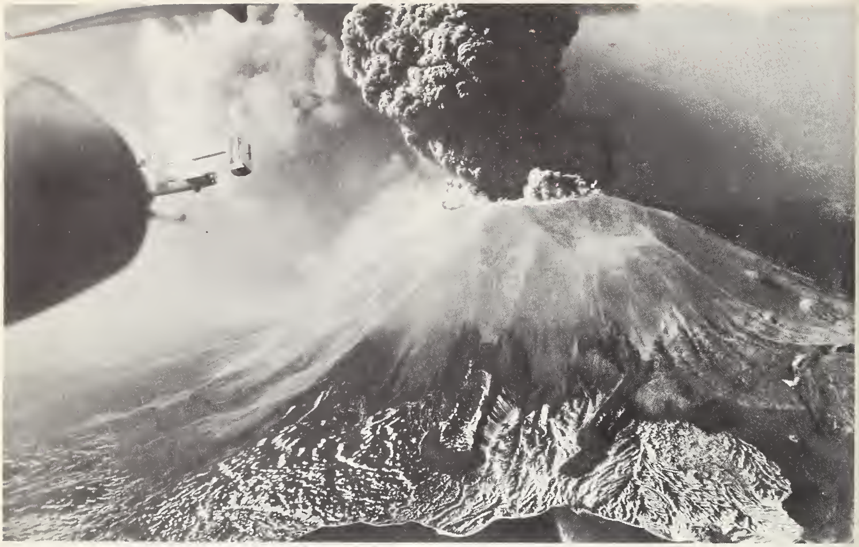


Fig. 8-11. Vesuvius in eruption.

is enormous, 400 miles to be exact; consequently, the slope from base to summit is very gentle.

Explosive Volcanoes. Granitic lava, as already stated, has a high melting point. It therefore readily solidifies, even in the crater of the volcano. Tremendous pressure builds up beneath the solid rock in the crater and, when the eruption occurs, liquid lava is blown high in the air, where, as tiny drops, it cools and hardens to form volcanic dust which may be carried in the air for hundreds of miles. The larger drops condense to solid particles, called ash and cinders, which fall near the crater. The viscous lava does not flow easily so that a steep-sided cone is formed, composed mainly of ash and cinders.

A well-known volcano of this type is Fujiyama, 60 miles southwest of Tokyo, Japan. This quiescent volcano (the last eruption was in 1707) has formed a cone of beautiful and symmetrical shape. To the Japanese, Fujiyama is a sacred mountain, and thousands of pilgrims make the ascent to the summit every summer.

The most familiar example of an explosive volcano is, of course, Vesuvius near Naples in Italy. Vesuvius has a tragic history. In the early days of the Christian era, Vesuvius was a supposedly extinct volcano. The soil at its base was very fertile, and the towns of Pompeii and Herculaneum grew and thrived in this pleasant land. Then, in the year 63 there was a violent earthquake in this region, and in 79 A.D. Vesuvius burst into violent activity. Pompeii was completely buried in ash and cinders, and Herculaneum was covered by mud. The fate of Herculaneum was due to the steam ejected from the volcano. The steam condensed to rain, and the rain brought down dust and ash as mud, enough to bury the city. Within the past century the volcanic debris has been removed from these ancient cities. In Pompeii the volcanic ash protected streets, buildings, and even paintings and mosaics, so that the destroyed city is now revealed in its former splendor.

Volcanoes in the United States. The only recently active volcano in the United States is Lassen Peak, near the northern

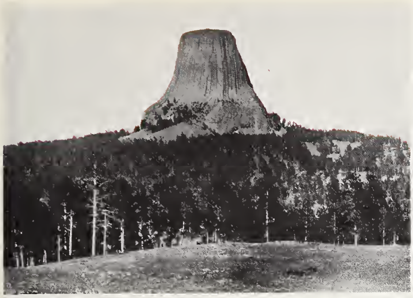


Fig. 8-12. Devil's Tower, a famous volcanic rock in Wyoming, 600 feet high.

tip of the Sierras in California. During its intermittent outbursts, Lassen Peak may be quiet or explosive; sometimes lava flows quietly down the slope of the cone and sometimes cinders are hurled high into the air.

There are, however, numerous cones of dormant volcanoes, the best-known being Mount Rainier (14,400 ft.) in Washington, Mount Hood (11,200 ft.) in Oregon, and Mount Shasta (14,000 ft.) in northern California. There are also many protruding necks of old volcanoes, particularly in the west. Devil's Tower in Wyoming is one of these. Devil's Tower was once the site of a magnificent volcano. When the volcano became extinct its crater was filled with solid basalt. Through the ages the rains have washed away the cone of cinders, and all that now remains of the volcanic mountain is a basalt column towering 600 feet above the countryside.

The Columbia Plateau is another reminder of intense volcanic activity in past ages. This vast area of more than 100,000 square miles in the states of Oregon, Idaho and Washington, has been covered by successive layers of basaltic lava. Since there are neither cones nor volcanic necks in this region, it is assumed that lava issued through cracks in the crust to form horizontal layers of lava—in places 5000 feet thick. Some of the basalt has weathered to form a fertile soil which yields rich crops of wheat and fruit.

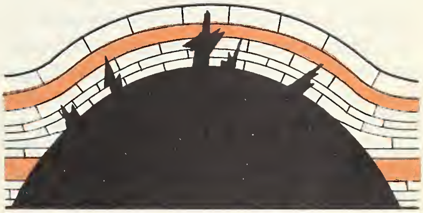
So far we have considered only igneous activity where molten lava has been ejected upon the earth's surface. In many igneous intrusions, however, molten lava never reaches the surface. Intrusions of this kind may form batholiths.

Batholiths. Suppose a crack in the crust does not extend all the way to the surface, what will the magma do in such a case? The magma may penetrate the rock layers, forcing the overlying rocks into the shape of a dome, and melting the lower layers of rock, as shown in Fig. 8-13. An intrusion of this kind is called a ***batholith***, a word from Greek meaning "stone of the deep," for batholiths do not rest on any other kind of rock, but extend deep into the earth. The Canadian Shield and the Coast Range contain many batholiths.

Batholiths sometimes cover areas of thousands of square miles, and form the cores of the great mountain ranges. In some regions of the Rockies and the Sierra Nevadas, for instance, the batholithic granite core has been exposed by the erosion of the overlying sedimentaries.

Old and Young Mountains. Fig. 8-15 is a cross section through Canada from Newfoundland to Vancouver Island along the 50th parallel of latitude. It includes old mountains in the east and young mountains in the west. What are the characteristics of old and young mountains? How do they differ?

Fig. 8-13. A laccolith is an intrusion of molten rock which forces the overlying rock into a dome-shaped peak.



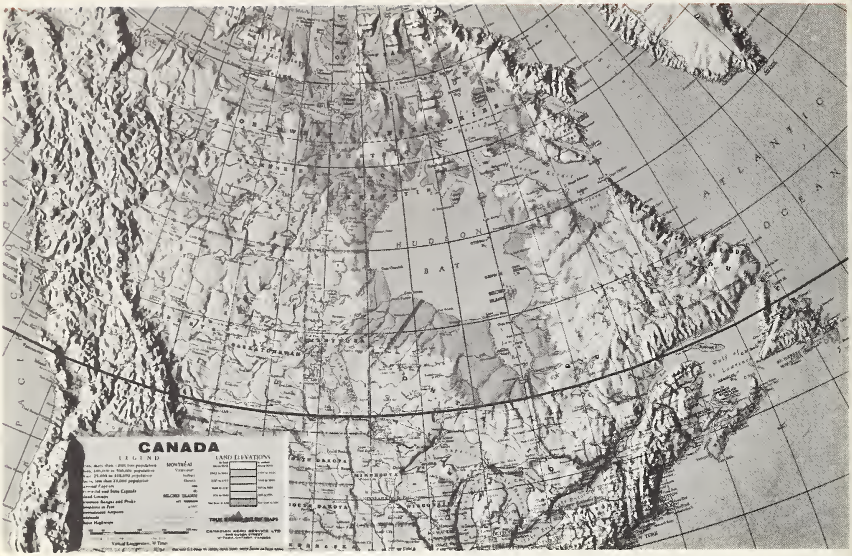


Fig. 8-14. Flying across Canada from Newfoundland to Vancouver Island along the 50th parallel of latitude, we would cross the Canadian Shield, the Great Plains, the Rockies, the high plateaus of central British Columbia, and the Coast Range.

In the west constructive forces dominate the destructive forces, so that the mountains are still growing. The growth of mountains is accompanied by earthquakes and volcanic eruptions. As we recall, both of these phenomena have occurred in the west in relatively recent times. Young growing mountains are immediately attacked by the destructive forces of weathering and erosion. As a result, they are high and rugged, their peaks are sharp, their slopes steep and barren, and their valleys narrow. The summits of the ridges of young folded mountains are also the tops of the anticlines, and these are most likely to be the first places stripped by erosion. The Rockies and the Sierra Nevadas are young mountains, but the youngest in Canada are the Cascades on the west coast. These were raised from the sea only about one million years ago. The intense volcanic activity which resulted in the deep basalt deposits in Washington, Oregon, and Idaho was a part of this same mountain-building process.

By contrast, in the east, the destructive forces dominate constructive forces. Here the mountains have stopped growing and erosion goes on unchecked. The Appalachians, the Adirondacks, and the Laurentian Mountains are old compared with the mountains west of the Great Plains. It is not surprising that, after a long period of erosion, their peaks have been lowered, their slopes are gentle, and their valleys broad. In short, the particular kind of mountainous scenery you see is determined by age. Youth in mountains is associated with bare, rugged peaks which tower above the snow line. But, in maturity, the sharp outlines have been rounded by weathering, and harsh scenery gives way to a more placid countryside.

Mountains of the Canadian Shield. The Canadian Shield forms a U-shaped area around Hudson Bay. The Shield is formed from the roots of the oldest mountains in Canada, which were worn down to a hummocky lowland and covered in

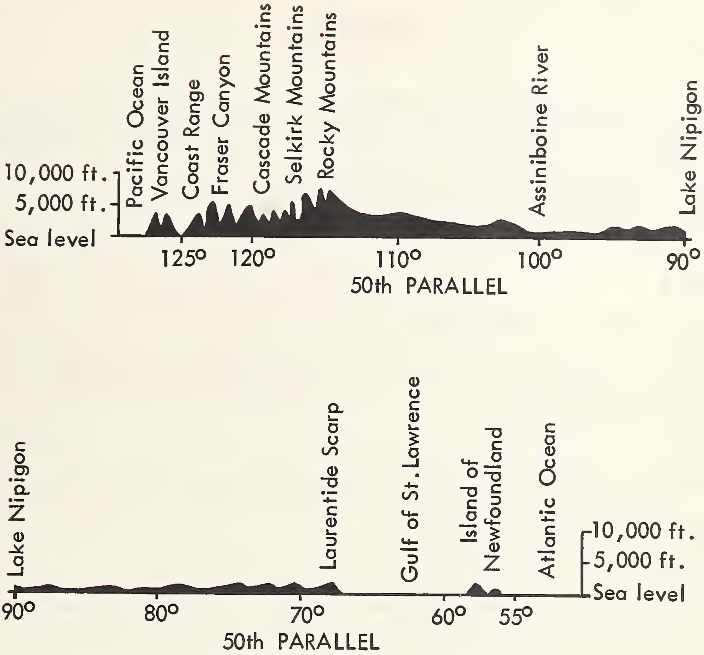


Fig. 8-15. A section through Canada along the 50th parallel of latitude.

part by sediment. This soil was later scraped away by the great continental ice sheets which spread over this region. The resulting terrain is rough and full of small lakes with many islands. The area now is not very high: approximately 1000 feet above sea level, increasing to about 5000 feet on the Labrador coast. Since the continental ice sheet melted, the Shield has been slowly rising, and is still rising today. If this continues, it may result in the building of a new mountain range on the Labrador coast.

Things to Remember

Mountains and plateaus are formed by the shrinking of the earth's interior.
The major mountain ranges of the earth are formed from ***geosynclines.***
There are ***three types of mountains:*** folded, block, and volcanic.
Anticlines are the upfolds of folded rock layers, ***synclines*** the downfolds.
A ***fault*** is formed when rock layers break and slide up or down along a break.
Volcanoes are of two kinds, quiet and explosive.
The only ***active volcano in the United States*** is Lassen Peak in the Sierras.
A ***batholith*** is an igneous intrusion which extends deep into the earth.

Questions

GROUP A

1. Name the three kinds of mountains and give an example of each.
2. What is (*a*) a syncline, (*b*) an anticline? Draw diagrams to illustrate these features.
3. Draw a diagram of a fault and explain what it means.
4. Give examples of (*a*) a quiet volcano, (*b*) an explosive volcano, (*c*) a dormant volcano. ([illegible] c ~ e

GROUP B

5. Explain how the cooling of the earth's interior causes the elevation of mountains.
6. What is a geosyncline? How did the Appalachian geosyncline give rise to the Appalachian Mountains?
7. How does isostasy tend to offset the erosion of a mountain peak? Draw a diagram to illustrate your answer.
8. What are block mountains and how are they formed? Give examples of block faults in the Rocky Mountains.
9. Why are some volcanoes quiet and others explosive?
10. What is a batholith? Explain.
11. Compare the mountains in Alberta and British Columbia with the mountains in Quebec and New Brunswick. How do you account for the marked differences?

Things to Do

Geological models · Photographs and diagrams frequently fail to illustrate the three-dimensional nature of various geological structures. This can best be done with modelling clay of various colors. Arrange layers of the colored clay into a block. This corresponds to the horizontal layers of sedimentary rock when first laid down.

This block can be used to demonstrate numerous features. First, slowly squeeze the block (at right angles to the layers) between two boards. This simulates folding, and a syncline and an anticline are formed. Then cut off the top of the anticline. This simulates the erosion of an anticline and shows the nature of the exposed rocks. Similarly a cut can be made through the strata and one of the sections raised or lowered to simulate a fault.

Block mountains, plateaus, and other features can readily be illustrated in this way.

9

THE RECORD OF ROCKS

THE HISTORY of the earth is written in stone. The rocks reveal not only the age of the earth, but they also show that the long history of the earth can be divided into natural periods. For example, the earth has an ancient history, a history of the middle ages, and a modern history, just as human history has. But let us return to the age of the earth. How is this determined? The answer to this question concerns radioactivity and may be difficult to comprehend at this time. However, we shall say more about this subject in Chapter 19.

Radioactive rocks contain the element uranium. Uranium atoms disintegrate to other atoms like radium and, finally, to lead. A lump of uranium changes one half of itself into lead every four and a half billion years. In other words, after 4.5 billion years, half of the atoms in a lump of uranium would have changed to lead; after another 4.5 billion years, half the remaining uranium would have changed to lead, and so on indefinitely. By finding the weights of uranium and lead in a given rock sample, scientists can calculate how long the radioactive change has been going on. And this, of course, is the approximate age of the rock. By examining radioactive rocks from many different parts of the earth, scientists have concluded that ***the solid earth was formed at least three billion years ago.***

Historical events, however, are revealed by studying sedimentary rocks. It is true that sedimentary rocks tell us only what happened on the continental shelf, but by inference we also gather information concerning geological events on the continental masses. The geologist first examines the strata, their composition and the sequence of the beds. Let us see why this information is so important to him.

Strata. If you look at sedimentary rocks in the side of a hill or quarry you may notice that the rocks are arranged in layers, or ***strata.*** How were the strata formed? They were, of course, formed under water. The site of the quarry—the place where you stand to observe the strata—was once covered by sea. Sediment from the land was carried by rivers and deposited at this place.

When a river deposits its load it does not dump all the sediment in one place; there is a sorting of sediments according to size. As the speed of the running water slackens, the first load to be dropped is gravel, which forms conglomerate; the next is sand, which forms sandstone; and, finally, far out to sea where the forward motion of the water is stopped, mud, which forms shale, is deposited.

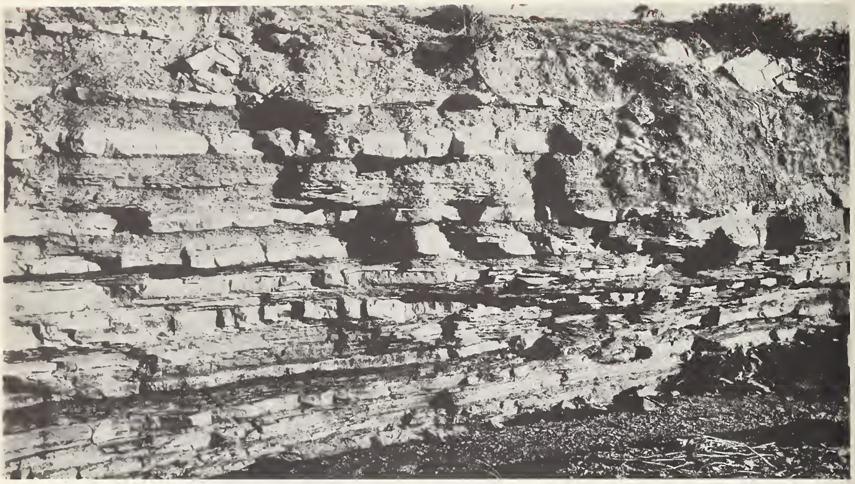
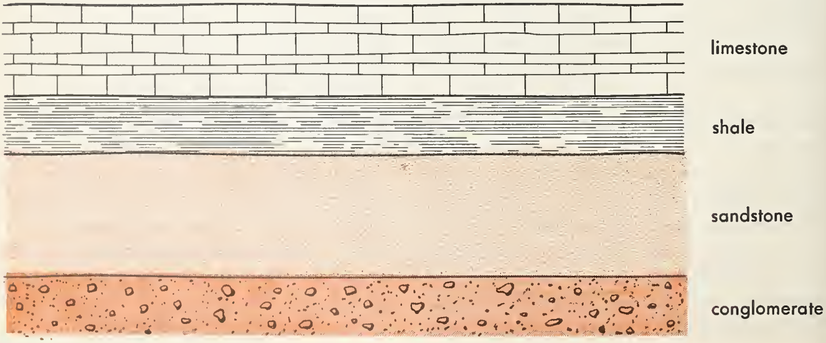


Fig. 9-1. Horizontal strata. Alternate beds of limestone and shale.

If we observe the strata more carefully we may notice that even a given layer of sandstone or shale shows a difference in texture; there may be finer and coarser deposits in the same stratum. Why is this? During spring rains the velocity of a river increases, and, as a result, coarser sediments are carried farther out to sea before they are deposited. As the speed of the river slackens finer sediments will be deposited on the coarser ones. Thus, depending on the speed of the river, each new deposit will have a different texture from that of the older deposit below. Differences in texture may be observed in thicknesses of less than a quarter of an inch, but the main strata themselves may be several feet thick.

The Sequence of Strata. The strata in the quarry follow a certain order or pattern. Let us suppose that the lowest layer is conglomerate, followed in turn by sandstone, shale and limestone, a sequence shown in Fig. 9-2. How could such a sequence of strata be interpreted geologically? What particular series of

Fig. 9-2. Horizontal strata. Can you explain the geological events that account for these particular sedimentary deposits?



events would account for this sequence?

Conglomerates are formed from very coarse deposits, carried by streams with high velocity. We can therefore reasonably assume that the streams were flowing from highlands, from mountains in fact. As, through the ages, the highlands are eroded, the velocity of the streams diminish and, in consequence, the sediments become finer. The sandstone layer therefore indicates a lower land mass than that which supplied the conglomerate, and the shale a still lower land mass. Finally, when the highlands are worn down to base level, that is, when the final stage of erosion is reached, rivers cannot flow at all, and no sediments are brought down from the land. The only sediments that now collect on the ocean floor are the fossilized remains of sea animals and plants which give rise to the limestone layer.

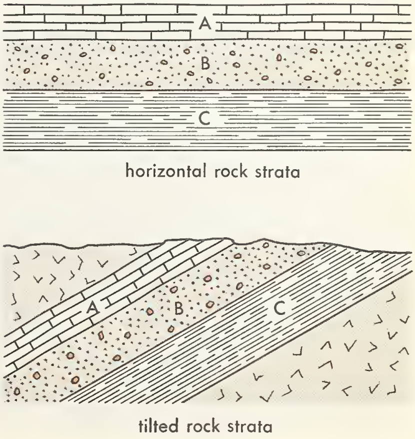
It should now be apparent that the nature of sedimentary deposits gives a picture of the gradual changes of nearby land. Suppose, for instance, that a deposit of shale is followed by a conglomerate instead of a limestone deposit as in the above example; that is to say, suppose the conglomerate rests upon shale. What would this indicate? The most likely inference to be drawn is that the neighboring land had been suddenly elevated. This would increase the speed of the streams and account for the coarser sediments of conglomerate.

In the same way, the sequence of the sedimentary deposits tells us something of the history of the Colorado Basin. In carving the canyon, the Colorado River first cut its way through about 4000 feet of sedimentary rocks consisting of red sandstone, white limestone, and gray-blue shale. Then it carved a gorge, about 1000 feet deep, into the igneous rocks below the shale. It is evident that for millions of years this region was covered by sea. And since sandstone rests upon limestone, which in turn rests upon shale, we can make some inferences concerning the configuration of the nearby land when these deposits were being laid down. What inferences can you draw from the above evidence?

Younger Beds Rest on Older Beds. In order to interpret the records of rocks, a geologist looks for clues just like a detective working on a criminal case. Some information is given by the nature and sequence of strata, more information is given by fossils, and still more by unconformities, a topic we shall discuss. Moreover, he compares the clues he finds in one region with those he finds in another. When he puts these pieces of evidence together, he has a fairly complete picture of the boundaries of the continents, the climatic changes in a given region throughout the ages, and of the development of life on earth.

One of the most obvious and important deductions concerning sedimentary rocks is that ***younger beds always rest upon older ones.*** This is true of all sedimentary rocks everywhere, and it has enabled geologists to compare the ages of rocks in widely different regions. For example, in the illustration of horizontal rock strata in Fig. 9-3 the uppermost layer, A, is the most recent one; the sec-

Fig. 9-3. Younger beds rest upon older ones. Layer A is more recent than layer B, and B is more recent than C.



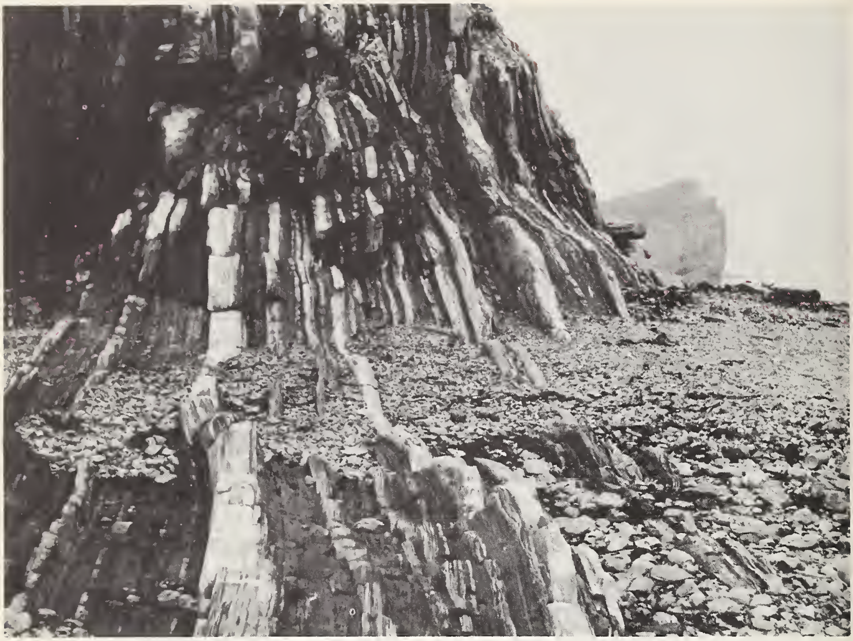


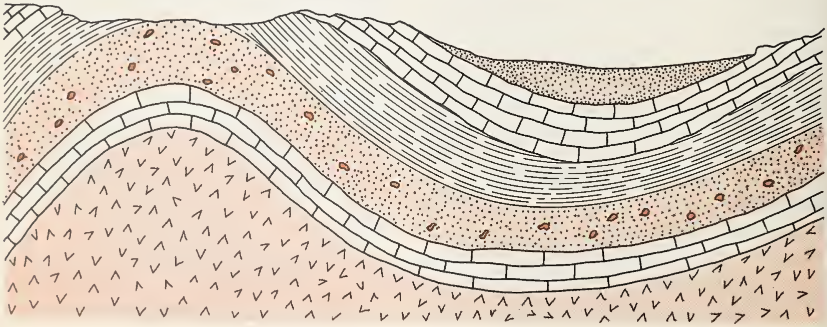
Fig. 9-4. Vertical strata. These beds were laid down horizontally, but they have been tilted through almost ninety degrees by earth movement.

ond layer, *B*, was deposited earlier than *A* but later than the lowest layer, *C*.

The other part of the same figure shows rock strata that have been tilted. Again, it should be evident that the bed on the extreme right is the oldest, and the bed on the extreme left the youngest. In other words, rock layers are like the pages of a diary—we can turn them backwards in time, from recent records to older ones and still older ones.

Unconformities. Let us imagine a particular region where sediments were laid down for millions of years. When, fol-

Fig. 9-5. Here rock strata have been folded and some of the surface rocks eroded.



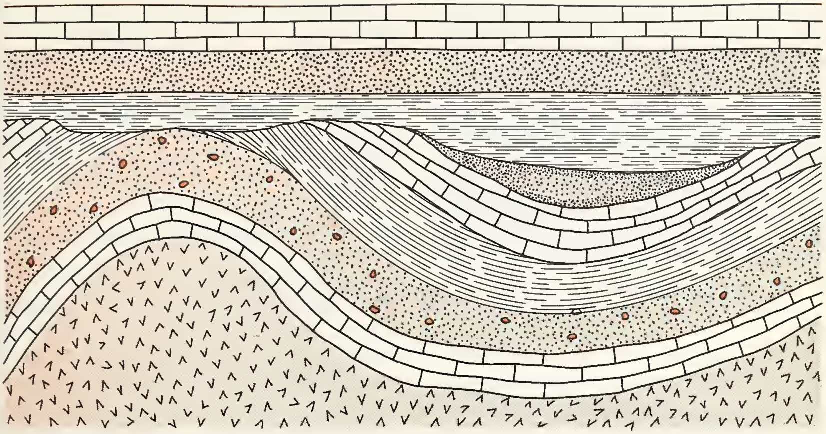
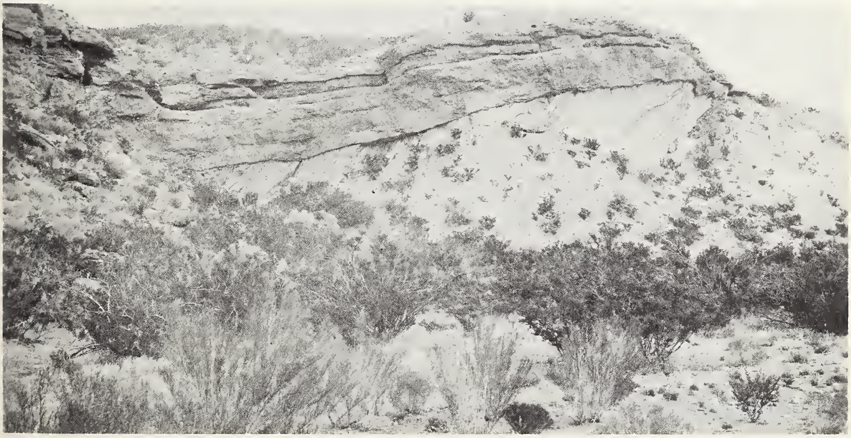


Fig. 9-6. An unconformity. The recent horizontal beds rest unconformably upon the older folded beds.

lowing the usual course of events, the sediments were elevated, the rock strata became tilted and folded. Above sea level, the exposed rocks were weathered and eroded, and in the course of time, the strata may have looked like those shown in Fig. 9-5. If, at a later period, a movement of the crust lowered these rocks below sea level, a fresh series of deposits would be laid on top of them as horizontal strata (see Fig. 9-6). There would be a surface of separation between the newer and older beds. Such a surface is called an ***unconformity.***

Unconformities are common. They are important clues to geological history because they indicate a long lapse of time between the lifting of the lower or older rocks and the formation of the upper or younger beds. Moreover, this lapse of

Fig. 9-7. The upper beds in this cliff are almost horizontal. The lower beds slope much more than the upper ones. The unconformity separating upper and lower beds is clearly visible. Notice that much of the lower beds is covered by talus.



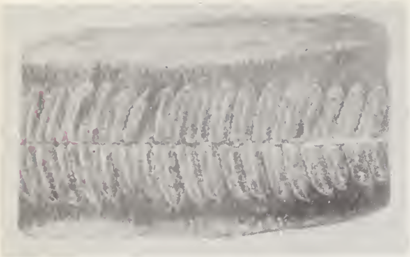


Fig. 9-8. Fossil fern embedded in shale. These ancient ferns lived in Carboniferous times.

time was, for the most part, a period of erosion when the region was dry land.

Fossils. If you examine a sedimentary rock you may see seashells embedded in it. These shells are the remains of animals that lived ages ago; they are ***fossils.*** The term ***fossil*** has several meanings. Some fossils are the shells or bones of animals; others are the impressions of shells left in rocks after the shells themselves have been dissolved; still others are the footprints of animals or the imprints of leaves. (See Fig. 9-8.)

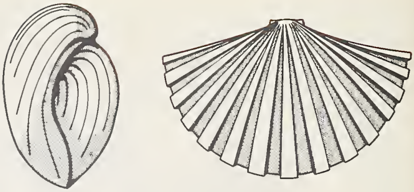


Fig. 9-9. Brachiopods are bivalve shells.

To be preserved as a fossil, the shell must be protected against the chemical effects of air or the dissolving action of water. Quick burial in sand or mud or ice is therefore favorable to fossil formation. For this reason, fossils of land animals are relatively scarce. In the sea, however, conditions are more favorable. When a sea animal dies, it sinks to the sea floor and the soft parts decompose. If the hard shell is covered by sediment before sea water has seriously affected it, the shell ultimately becomes a fossil embedded in sedimentary rock.

The most common of all fossils are the ***brachiopods;*** they are among the

Fig. 9-10. One of the richest deposits of fossil bones of dinosaurs, extinct reptiles described on page 94, is in the Red Deer River valley in Alberta. Here, in the Dinosaur Provincial Park northeast of Brooks, a palaeontologist is uncovering the skeleton of a duck-billed dinosaur for a field exhibit.

earliest known fossils and have survived to the present day. The body of a brachiopod is enclosed in a shell which consists of two valves or parts somewhat like a clam shell. It is, of course, only the hard shell of the brachiopod that forms a fossil. Another animal that has lived from ancient times to the present day is the ***coral***. Corals have always thrived in a warm climate, and their fossils have made reefs in many parts of the world.

The long time interval represented by an unconformity has a marked effect upon the fossils of a given species. By comparing the fossils of, say, brachiopods above and below an unconformity, we learn that the animal has changed or evolved during the lapse of time represented by the unconformity. Sometimes the animals die out completely, so that some species are not found in the beds above an unconformity.

Things to Remember

The ***age of the earth*** is probably at least 3 billion years.

The ***younger beds*** of sedimentary rocks always rest upon ***older ones.***

An ***unconformity*** is a surface of separation between newer and older beds.

Fossils are bones or shells or impressions of animals and plants that lived in past ages.

The ***age of sedimentary strata*** can be found from the fossils they contain.

Questions

GROUP A

1. Fossils are never found in igneous or metamorphic rocks. Why not?
2. Why are there no rock strata to represent the time lapse represented by an unconformity?

GROUP B

3. Explain how the age of a radioactive rock can be found.
4. (*a*) What is an unconformity? Draw a diagram. (*b*) How was it formed and what does it indicate?
5. Explain how a geologist gets information concerning the relative ages of rocks.
6. The following is the rock sequence of the Grand Canyon: (*a*) sandstone, (*b*) limestone, (*c*) shale, (*d*) igneous rock. What can you conclude regarding the changes in the land mass of the Colorado Basin?
7. Discuss the conditions necessary for the formation of fossils.

Things to Do

A fossil collection · Collecting fossils is most rewarding, and even exciting. It can be a most interesting hobby even through adult years. Write to the nearest college or university geology department and inquire if there are rocks in your neighborhood likely to yield fossils. Ask also for information concerning the kind of fossils to look for. Then, if possible, visit a museum where these fossils are on display. Then take a hammer, go to a cliff or quarry, and begin the search. Your first fossil will be an exciting discovery. Does your fossil belong to an extinct species? How many million years since the quarry was under water? More will be said of the significance of fossils in the next chapter.

10

EARTH HISTORY

Many times through the ages great floods, which lasted for millions of years, have covered large portions of the earth. During the periods of submergence, deposits were laid down and fossils collected in the beds. Flooding was followed by an elevation of the land, and the forces of erosion then came into action. When the land was again submerged, an unconformity was formed between the new and the old beds. The fossils in these new deposits show that the animals of the older beds have changed greatly in structure, and some have even become extinct.

By means of unconformities, fossils, and other evidence, geologic time has been divided into five great ages called ***eras.*** Actually the eras are separated from each other by a ***revolution.*** The term ***revolution*** means, geologically speaking, a period of widespread mountain building which resulted in marked changes on earth, climatic and otherwise, so that life before and after a revolution was markedly different. Each of the eras is named according to the type of life that existed at that time. These are (1) Archeozoic, a word meaning ***earliest*** life; (2) Proterozoic, a word meaning ***primitive*** life; (3) Paleozoic, ***ancient*** life; (4) Mesozoic, ***middle*** life (5) Cenozoic, ***recent*** life. To these five could be added a sixth, the Azoic era, a word meaning ***no life.*** The Azoic era would refer to the long epoch when the crust of the earth was cooling and hardening, the long span before plant or animal life developed on earth. The three eras, the Proterozoic, the Archeozoic, and the Azoic, are often combined and referred to as the Pre-Cambrian era.

Within each era there are ***periods*** when the sea inundated large areas of land. Towards the end of a period the land was again uplifted so that it emerged from the sea. Period changes were neither as extensive nor as prolonged as those which marked the end of an era. Nevertheless, they are important geologic landmarks. The comparative lengths of the eras and periods are shown on page 91. Notice that seven periods make up the Paleozoic era, and three periods make up the Mesozoic era. Man's existence dates from the latter part of the Cenozoic era, an era of comparatively short duration—about 60 million years, culminating with the present time.

The Archeozoic Era. The first evidence of life is found in the Archeozoic era. The evidence is slim and unreliable however; it consists merely of graphite deposits found in some of the sedimentaries. Since graphite is a form of carbon, an element essential to plant life, these deposits seem to indicate that there was some form of plant life in the sea at this time.

The rocks for the most part are igneous and metamorphic in nature, actually granite and gneiss. The largest outcrop of Archeozoic rock is in Canada, where about two million square miles of gneiss are exposed. This extensive deposit indicates that the thin crust of the earth probably cracked wide open so that vast amounts of molten magma poured forth from its interior. Erosion of the igneous rock then took place for hundreds of millions of years. It is not, therefore, surprising that sedimentary deposits of great thickness were laid down. Through the ages these sedimentaries have become folded, crushed, and injected with igneous intrusions.

At the close of the Archeozoic era the so-called Laurentian revolution took place. This was a period of intense mountain building, particularly in Canada where the vast Canadian Shield, as it is called, was raised. The Canadian Shield is an exposed area, fabulously rich in minerals. Through the ages it has kept its head above water although the region around this continental shield has been flooded a score of times by shallow waters.

The Proterozoic Era. The uplifting process of the earlier era extended the boundaries of North America so that during the Proterozoic era it was considerably larger than it is today. Life in the seas was more plentiful than in Archeozoic times. At least there are thick deposits of graphite in the slates of this period, and the simplest way to account for this is to as-

Era	Approximate Length	Period	Approximate Length
Cenozoic	60 million	*Pleistocene* *Pliocene* *Miocene* *Oligocene* *Eocene*	2 million 10 million 15 million 10 million 20 million
Mesozoic	120 million	*Cretaceous* *Jurassic* *Triassic*	60 million 30 million 30 million
Paleozoic	340 million	*Permian* *Carboniferous* *Devonian* *Silurian* *Ordovician* *Cambrian*	40 million 70 million 40 million 30 million 60 million 100 million
Proterozoic	500 million		
Archeozoic	500 million		
Azoic	2500 million		

sume that the carbon is derived from marine organisms, probably algae. Although no fossils have been definitely recognized in the sedimentary deposits of this era, there is reason to believe that the seas abounded in worms and other simple forms of life.

The era closed with the elevation of high mountains. Then followed an unusually long period of erosion before the beginning of the Paleozoic era.

The Paleozoic Era. As might be expected, many of the "pages" of stone are missing from the ancient diary of the Archeozoic and Proterozoic eras; they were destroyed in the upheavals of those long periods. In the Paleozoic era, however, the record is more detailed and, as a result, the chronological sequence is more complete.

At the beginning of the era, the North American continent was uplifted beyond its present boundaries, and land extended for hundreds of miles into the Atlantic and Pacific oceans. At a later period, water almost completely covered the continent for millions of years. Sediments of these early Paleozoic seas have various uses in industry today. For example, building stone was formed in this era in eastern Canada and Manitoba, and minerals such as the iron ore in Newfoundland were also laid down then. The extensive salt and potassium chloride deposits that underlie the Great Plains were formed from the dried-up arms of the sea trapped by the rising land. Much of the oil and natural gas in Alberta is found in Paleozoic sedimentaries.

Many outcrops of rock of this era occur across Canada. Mount Robson and Mount Stephen in British Columbia, and the mountains around Lake Louise, all contain rocks of the Cambrian period. The limestone escarpment that forms Niagara Falls is of the Silurian period.

Early Paleozoic Life. Paleozoic deposits are remarkable because they contain the earliest fossils, fossils that appear suddenly and in abundance. How can this be explained? It is supposed that many kinds of soft-bodied animals existed before the Paleozoic era. These animals found food and shelter among the seaweeds that were plentiful. As the animals increased in numbers there was more and more competition for the limited supply of food and, in consequence, they probably began to attack and devour each other. Then, in the course of time, they learned how to use calcium carbonate extracted from sea water to form hard shells as a protection against enemies. It was the invention of armor by sea animals that made possible these earliest definite records in sedimentary rocks.

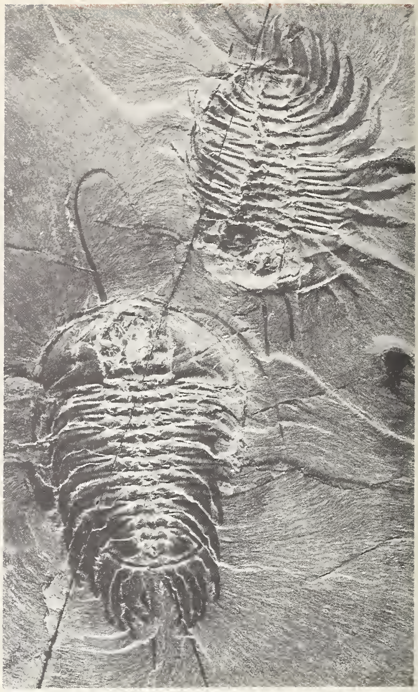


Fig. 10-1. Trilobites in shale. Note the antennae and legs of these animals.

The most common fossil of this period is the trilobite, a lobster-like animal which was divided into three lobes, or divisions by two furrows extending the whole length of its body. (See Fig. 10-1.) They were small animals, usually less than four inches long, although some specimens

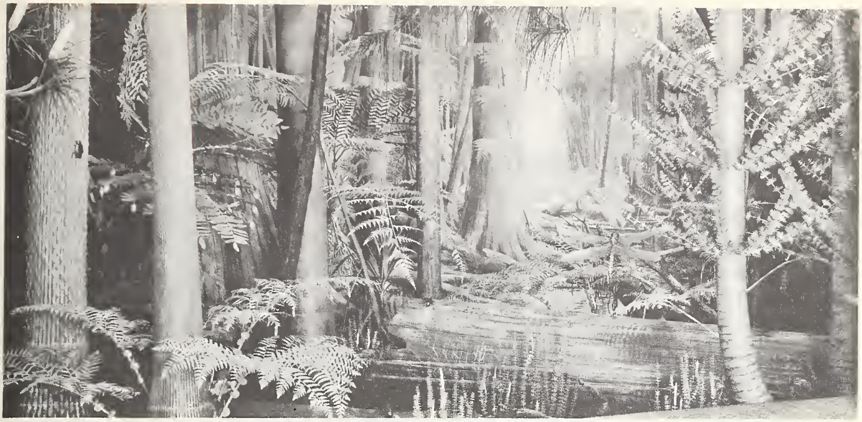


Fig. 10-2. An artist's concept of a Carboniferous forest. Notice the fern trees and horsetails.

grew to at least a foot in length. Trilobites were far more numerous than all other animals in the Paleozoic seas. They flourished for millions of years and then, for some unknown reason, became extinct at the close of the Paleozoic era.

The Late Paleozoic Era. During the Paleozoic era mountains were raised in North America and Europe, including the mountains of New Brunswick, Vermont, and New Hampshire. This mountain building was followed by two geological periods known as the Mississippian and Pennsylvanian, or, collectively, the Carboniferous period or the Age of Coal.

During the Mississippian, the central part of North America was covered by sea; the famous limestones of the Mississippi valley were laid down, particularly the Indiana limestones which are amongst the best building stones anywhere. The Pennsylvanian period began with the receding of the seas from central North America. This region became swampy: vegetation thrived and ultimately gave rise to extensive coal deposits. The forests of this period bore little resemblance to our modern forests. There were no trees with leaves such as maple, beech, or oak, nor were there flowers or fruits. Instead, there were dense forests of club mosses, horsetails and ferns that grew into gigantic trees 100 feet in height and with trunk diameters of four feet or so. The coal fields of New Brunswick and Nova Scotia were formed at this time, as were most of the coal fields in Europe and other parts of the world.

For millions of years during the early Paleozoic, the only living animals were ***invertebrates,*** animals without backbones. In the middle of the era fishes appeared for the first time; they were the first animals with backbones, usually called ***vertebrates.*** The great variety of fishes in the seas today have developed from these few strange Paleozoic fishes. Then, in the latter part of the era, came the amphibians, descendants of fish. Amphibians lay their eggs in water and the embryo, equipped with gills like a tadpole, hatches in water. At a later stage in life, a lung replaces the gills, and limbs appear so that the creature can live on land. Amphibians grew in size and flourished; then, for some reason, they almost became extinct in Mesozoic times and, today, the only remaining amphibians are a few small species such as frogs and salamanders.

The Paleozoic era ended in the great Appalachian revolution, a period of intense mountain building all over the world. This was the time when the original or ancestral Appalachians were

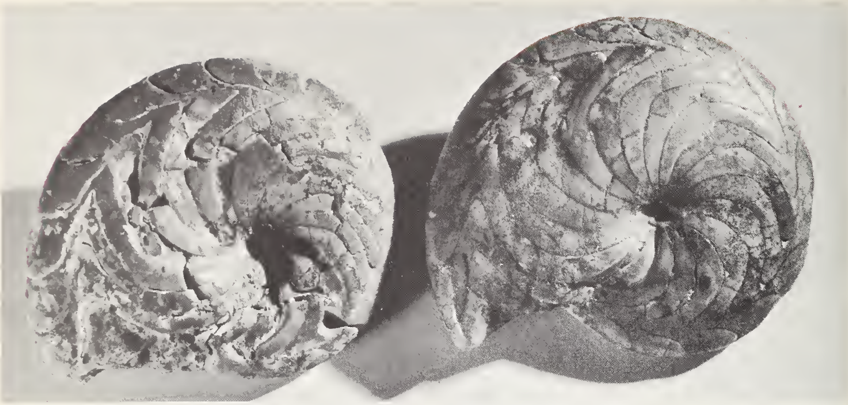


Fig. 10-3. Ammonites are coiled like a watch spring.

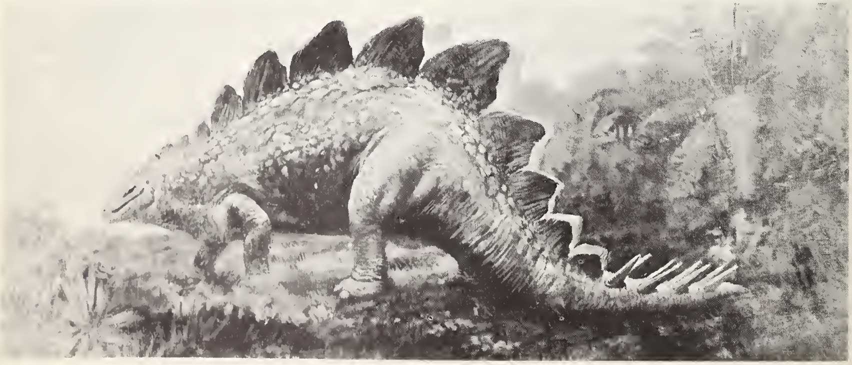
raised, a magnificent range of mountains five miles high.

The Mesozoic Era. The Mesozoic era is a period of transition between ancient and modern life; it corresponds to the Middle Ages in human history. Many changes resulted from the Appalachian revolution at the close of the Paleozoic: lowlands were elevated so that, in general, the climate became colder; entire species of plants and animals were wiped out in the new environment; and organisms that survived underwent vital changes.

The most numerous shellfish were ***ammonites***. Their fossils are easily recognized because the shells were coiled like a watchspring. However, the ammonites died out in the revolution at the end of the Mesozoic. An ammonite fossil therefore dates a rock as Mesozoic, just as a trilobite dates the rock in which it is found as Paleozoic.

The Mesozoic is characterized by the development of life on dry land. The amphibians of the Paleozoic developed into reptiles, giant monsters called dinosaurs, which conquered the earth and remained in unchallenged possession for at least a

Fig. 10-4. This member of the dinosaur group (stegosaurus) was armed with a double row of plates and spines down his back (restoration by Charles R. Knight).



million years. The dinosaurs were probably the largest, the most ferocious and powerful land animals of all time (some weighed as much as 50 tons) and yet they, too, came to an abrupt end at the close of the Mesozoic. Today the only survivors of these mighty reptiles are snakes and lizards.

Some of the smaller reptiles took to the air during the Mesozoic and became birds, birds with long tails and sharp teeth. The ability to fly was another advance in the life process as we see it unfolding; first from sea to land animals, then from land animals to animals that fly.

During the latter part of the Mesozoic, shallow seas inundated the Rocky Mountain region. But towards the end of the period, the land in this region began rising again, and the era closed with another characteristic revolution accompanied by much volcanic activity and mountain building. The Rocky Mountains were uplifted at this time. The original Appalachians, elevated at the close of the Paleozoic, were eroded to a plain during the Mesozoic. The remaining synclinal folds of the ancestral Appalachians were now re-elevated to form the present Appalachian Mountains. By the end of the Mesozoic the continent of North America had assumed the same approximate shape and topography that it has today.

Most of the rocks that are exposed in the western plains are of the Cretaceous period or of the early Cenozoic era. (One can find fossils of dinosaurs in many places where these rocks are exposed.) The Athabasca tar sands and the coal seams of Alberta were formed in this period.

The Cenozoic Era. The Cenozoic era began only about 60 million years ago; it is the era in which we are now living. Most of North America has been above sea level throughout the Cenozoic. However, the sea remained over the Gulf states for some millions of years, during which time sandstones and limestones were deposited in Texas and Louisiana. These formations contain enormous accumulations of sulfur, oil and natural gas, and are of great economic importance; about one-half of the oil drilled in the United States is obtained from these Cenozoic deposits.

The Cypress Hills, and the Swan Hills of Alberta, consist of conglomerate, marl (white calcium carbonate which was deposited at the bottom of a lake), and sands of the early Cenozoic era.

About the middle of the era there was

Fig. 10-5. The woolly mammoth was common in Eurasia and North America some 25,000 years ago. It is now extinct.

intense earth movement in the west accompanied by volcanic activity. The Sierra Nevadas were raised at this time and also the Coast Range in California. Repeated lava flows that covered parts of Oregon, Washington, and Idaho to a depth of almost a mile were a part of this same crustal violence.

The Cenozoic is often called the ***Age of Mammals.*** Mammals (animals that give milk to their young) date back to the Mesozoic but they were no larger than a cat or a small dog, and no match for the giant reptiles. When reptiles became extinct, the mammals developed rapidly in size and kind; they became the rulers of the earth, a distinction they hold to this day.

The most interesting sediments in the Cenozoic are those deposited by ice within the past two million years. There have been four glacial periods, alternating with warm intervals as the ice receded. It was during the warm interglacial periods, about one million years ago, that man made his first appearance. As stated on page 65, the maximum advance of the last ice age occurred about 25,000 years ago. Today we are still probably in its thawing stages. You will recall that, during this glacial period, the mountains of New England and the adjoining regions of Canada were completely covered by a continental glacier and that the Rockies and the Coast Range were scratched and eroded by valley glaciers.

Life and Time. Animal life on earth has developed from simple to more complex creatures. First the invertebrates, then the vertebrates. The earliest vertebrates were fishes, then came amphibians, then reptiles, then birds, then mammals, and finally man, the only animal with the ability to reason. Living creatures have always had the ability to adapt themselves to changing conditions. Even such lowly creatures as trilobites and ammonites had this ability. As their fossils are traced from older to more recent rocks, marked changes in structure are revealed. These changes helped the animal to move more freely so as to escape enemies, or to obtain food more easily. This response to changing conditions is what the geologist means by ***evolution.***

We must remember, however, that even when the simplest animal life first appeared the earth was already very old. Before this, for millions of years, there had been an elementary form of plant life, first in the sea and later on the land; and before this there was no life at all.

How does the history of mankind compare with the overall history of the earth? The incredibly short time occupied by human history is best illustrated by analogies used by well-known scientists. Professor Jeans, the astronomer, made the following interesting comparison: If the age of the earth is represented by the Empire State Building in New York (1250 feet), then the whole of recorded human history can be represented by the thickness of a dime placed on top of its tower. An equally striking analogy is made by Professor Dorn, a German scientist. Suppose the age of the earth (three billion years) is compared with a 24-hour-day, measured from midnight to midnight. If timed on the 24-hour clock, the beginning of organic life on earth occurs at 10:50 P.M., man makes his first appearance just 22 seconds before midnight, and all the events of world history, from 3000 B.C. to the present day, are crowded into the last half second before midnight.

Things to Remember

The ***six geologic eras*** are (1) Cenozoic (recent life), (2) Mesozoic (middle life), (3) Paleozoic (ancient life), (4) Proterozoic (primitive life), (5) Archeozoic (earliest life), and (6) Azoic (no life).

Geologic eras are subdivided into ***periods.***

The ***earliest recognizable fossils*** are found in early Paleozoic rocks.

Only ***invertebrates*** lived in early Paleozoic times; ***vertebrates*** appeared in the late Paleozoic.

The ***earliest vertebrates*** were fishes.

The Carboniferous period was the ***Age of Coal.***

The Mesozoic was the ***Age of Reptiles.***

The Cenozoic was the ***Age of Mammals.***

The North American continent had ***four glacial periods*** during the Cenozoic.

Questions

GROUP A

1. Explain the following terms and give an example of each: (*a*) invertebrate, (*b*) vertebrate, (*c*) amphibian.

2. Which of the following fossils are impôrtant in dating the era when a sedimentary rock was formed: (*a*) trilobite, (*b*) coral, (*c*) brachiopod, (*d*) ammonite? Why?

3. State whether the following are alive or extinct: (*a*) mammal, (*b*) trilobite, (*c*) dinosaur, (*d*) coral, (*e*) ammonite, (*f*) amphibian.

4. (*a*) What is meant by the term ***revolution,*** geologically speaking? (*b*) What is meant by the terms ***era*** and ***period?***

5. What is the evidence of life in the Archeozoic and Proterozoic eras?

6. When were the following mountains elevated: (*a*) the original Appalachians, (*b*) the present Appalachians, (*c*) the Rockies, (*d*) the Sierra Nevadas?

GROUP B

7. Fossils suddenly appear in abundance in the early Paleozoic. How can you explain this?

8. Discuss the nature of the vegetation that gave rise to the coal deposits.

9. Compare animal life in the Paleozoic and Mesozoic eras.

10. (*a*) Briefly describe two major geological events in the Cenozoic era. (*b*) Name two fossils you would expect to find in Cenozoic rocks, and two you would not expect to find.

11. State briefly what you know of the development of animal life as revealed by rocks.

Things to Do

A geological time chart · Draw a line 30 cm. long to represent geological time (3 billion years). On this scale, one cm. represents 100,000,000 years, and one mm. represents 10,000,000 years. Divide the line according to the lengths of the geologic eras as shown on page 91. From this same table get the necessary information to mark such important geological dates as (*a*) the earliest recognizable fossils, (*b*) the earliest (i) amphibians, (ii) birds, (iii) reptiles, (iv) mammals, (*c*) the appearance of man, (*d*) the elevation of (i) the Rockies, (ii) the Appalachians, (iii) the Canadian Shield, (iv) the Coast Range, (e) the ice ages.

Weather

LOOKING AHEAD

The three air masses that determine most of the weather in Canada originate over the Pacific Ocean, in the Gulf of Mexico, and in northwestern Canada. The air mass from the Pacific is warm and moist when it reaches the British Columbia coast line, then becomes dry in its passage over the mountain ranges. The tropical air mass from the Gulf of Mexico is warm and moist. The polar air mass from northwestern Canada is cold and dry. Converging winds sometimes cause two of the air masses to meet, forming a front. If the polar air mass and the tropical air mass meet, the warm air is pushed "uphill" over the cold air and into the upper atmosphere. The rising air expands and cools so that moisture carried from the Gulf of Mexico condenses as rain.

Another familiar weather feature is wind. The winds most frequently in the newspaper headlines are tornadoes and hurricanes. Every year these violent storms cause a great deal of damage: tornadoes, usually in the midwestern United States in the spring; and hurricanes, usually along the east coast of North America in the late summer. Tornadoes sometimes form in July and August in southwestern Ontario and in the southern parts of the prairie provinces. Hurricanes moving along the Atlantic coast bring high winds into the Atlantic Provinces in August and September. The air in both tornadoes and hurricanes rotates counterclockwise in the Northern Hemisphere. A hurricane originates over tropical waters. It has winds of over 100 miles an hour and a diameter of 200 miles or so. A tornado originates over land. It has lashing winds of at least 300 miles an hour concentrated in a funnel less than a mile in diameter.

Man can do little to lessen the impact of these destructive natural forces, but the Meteorological Branch of the Department of Transport, a government agency, is alert at all times to give warning of approaching storms: hurricanes, tornadoes, thunderstorms, dust storms, sleet, and blizzards. The Meteorological Branch operates more than 300 weather stations, and its staff performs numerous services of particular use to aviation, agriculture, and industry. It is probably best known to the general public through its weather bulletins and forecasts which are issued every day.

← *Lightning flashes in a thunderstorm remind man of his inability to curb some forces of nature.* (Ewing Galloway)

11

WINDS AND AIR MASSES

MAN HAS always been at the mercy of the weather. Even scientists have been unable to control it and, today, weather is the result of natural forces just as it was centuries ago. From earliest times weather has affected man's food supply, his safety, and his comfort. Frosts, heavy rains, hail, wind, and prolonged droughts damage crops, but sunshine and gentle rain insure abundant food.

Fig. 11-1. RCAF pilots receive a weather forecast before starting a long flight.

We must not think that scientists have achieved nothing in their efforts to conquer the "elements." Far from it. Air conditioning, at least, makes man independent of weather in limited areas and, as we well know, travel in train, buses, and even in automobiles is made comfortable by air conditioning. Can you think of other examples where scientists have modified the weather? One that comes to mind is the dispersing of fog over airport runways and on main highways where traffic is heavy.

Before scientists can control weather they must know exactly what causes it. Meteorologists have flown through the heart of violent storms to learn what goes on there, and, dropping pellets of dry ice into clouds, they have produced rain. They know that weather is due to such factors as the unequal heating of the earth by the sun's radiation, the amount of water vapor in the air, and the rotation of the earth. But there are many aspects of the weather that are not completely understood. Until they are known, any large-scale control of the weather and climate will be impossible.

Although weather cannot be controlled, meteorologists are able to predict with considerable accuracy the weather for the following day or two. A timely prediction of bad weather gives people time to protect crops and property and warns aircraft so that they can by-pass storm areas. The benefits of accurate weather forecasting are so widely recognized that every civilized country has a weather-forecasting service. The Meteorological Branch of the Department of Transport, Government of Canada, provides this valuable service for our country and cooperates with the weather services of other nations as well.

Before discussing weather forecasting, let us consider some of the factors that account for the weather.

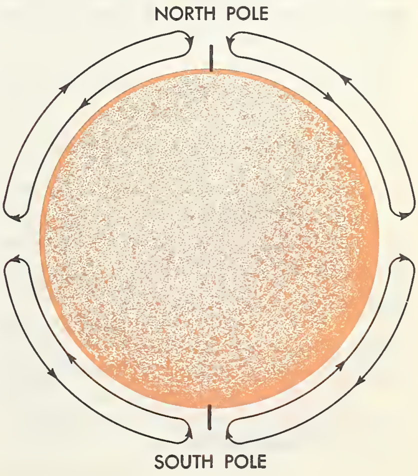
WINDS

Most of the heat radiation that the earth receives from the sun is absorbed at the surface. Different parts of the earth absorb differing amounts of heat; the Equator, for example, receives more heat than the polar regions. This unequal heating causes large-scale convection currents in the atmosphere.

We might expect the air above the Equator to rise vertically and flow, at high altitudes, directly toward the cooler regions north and south, and cooler, denser air from the North and South poles to flow close to the ground directly toward the Equator as shown in Fig. 11-2. If this were so, the surface winds in the Northern Hemisphere would always be north winds, that is, they would blow from the north; in the Southern Hemisphere they would be south winds, that is, they would blow from the south; and at the Equator and the poles there would be little or no wind.

Factors Affecting Winds. The circula-

Fig. 11-2 If the earth did not rotate, the Northern Hemisphere would experience north winds and the Southern Hemisphere would experience south winds at the surface of the earth.



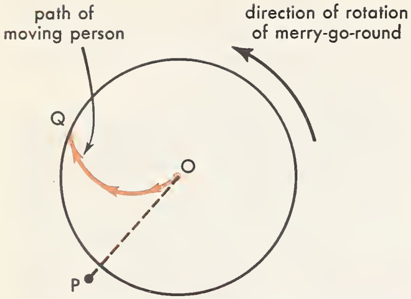
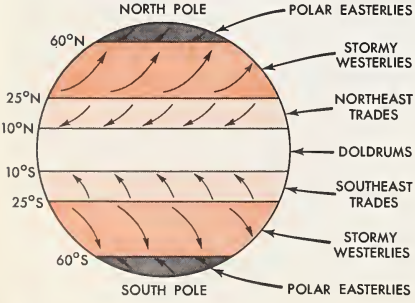


Fig. 11-3. If one walks straight out from the center of a rotating merry-go-round, one follows the curved path OQ over the surface of the merry-go-round.

tion of air about the earth, however, is not as simple as the above picture indicates. It is complicated by three major factors: (1) the rotation of the earth, (2) the irregular distribution of large areas of land and water, and (3) the season of the year. The rotation of the earth from west to east causes the winds to veer, and blow, not from the north or south, but from the northeast or southeast. Let us see why. Suppose we try to walk from the center to the edge of a merry-go-round while it is in motion; that is, suppose we set out to walk from O to *P* as in Fig. 11-3. A force due to the motion of the merry-go-round deflects us from our path so that we actually reach the point *Q* at the edge. That is, we have been deflected in a clockwise direction.

A similar force acts on the air as it moves from the poles to the Equator. In the Northern Hemisphere the deflection is clockwise, so that the northerly wind becomes a northeast wind; in the Southern Hemisphere the deflection is counterclockwise, so that the southerly wind becomes a southeast wind. The result is a circulation like that in Fig. 11-4.

Fig. 11-4. The wind belts of the world.



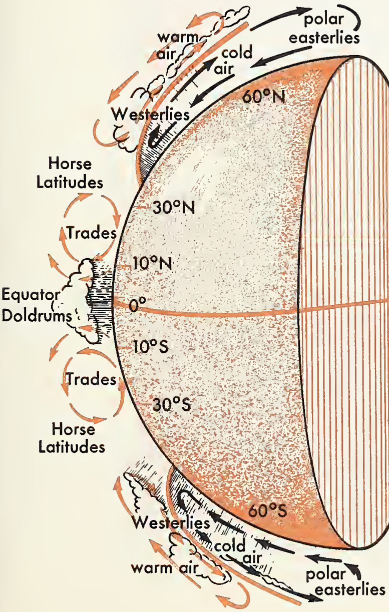
Prevailing Winds. What is wind? It is a horizontal movement of air over the earth's surface caused by a difference in pressure. If you were to observe and record the direction of the wind at a high and exposed place several times a day for several months, you would find that the wind blows more often from one direction than from any other. The wind from that particular direction is called the ***prevailing wind.*** In southern Canada, and also most of the United States, the winds blow generally from the southwest. These winds are known as ***prevailing westerlies.*** In the Southern Hemisphere over corresponding latitudes, the prevailing westerlies blow from the northwest.

You will notice in Fig. 11-4 that on both sides of the Equator, between latitudes 10° N. and 10° S., there is a region called the ***doldrums.*** Here there is little or no wind. In the days of sailing ships navigators tried to avoid this region. On a westward voyage they preferred to sail between latitudes 10° N. and 25° N., or between latitudes 10° S. and 25° S. Why? In these regions winds blow fairly steadily day and night from the northeast in the Northern Hemisphere and from the southeast in the Southern Hemisphere. As a result, ships sailing westward from Europe to North America found it advantageous to sail a route where the northeast trades would speed their ship westward. It was because these winds were so much help to trading vessels that they were named ***trade winds.***

Like the captains of sailing ships, pilots of aircraft take the direction of prevailing winds into account when planning long flights. More time is required to fly a plane from Halifax to Vancouver than is required to fly the same plane from Vancouver to Halifax. In addition to the longer time required, a larger supply of fuel must be carried on the east-west trip, with a resulting loss in payload. How can you explain this? It should now be apparent why records for trans-Atlantic and transcontinental flights have usually been made on west-to-east trips.

Now let us examine Fig. 11-5 and observe the vertical convection currents at different latitudes. Near the Equator, where the earth receives its greatest amount of heat, the heated air rises and there is little horizontal movement of surface air directly beneath this rising current. Just north and south of the doldrums, cool air moves in along the surface to replace the hot air rising over the Equator. This cool air, moving horizontally along the earth's surface constitutes the trade winds. Around the North and South poles there is little horizontal wind because here cold air is descending.

Fig. 11-5. The major vertical convection currents over the earth's surface. Compare this diagram with Fig. 11-4.



Winds and Atmospheric Pressure. Although Figs. 11-4 and 11-5 are more accurate and detailed representations of the prevailing winds around the world than is Fig. 11-2, they are still too simplified for accuracy. They take into account neither seasonal changes nor irregularities due to unequal heating of large land masses and oceans. A better picture of the world's prevailing wind systems is shown by the maps in Fig. 11-6, which show that the winds of the world are quite complex and differ in summer and winter. Note particularly that the winds in summer blow in the opposite direction from those in winter over east and southeast Asia. This reversal of the prevailing wind with the seasons is called a ***monsoon*** effect. It has a great deal to do with the climate of Asia and is more fully discussed in Chapter 15. Notice also that the east coast of North America experiences a similar monsoon effect.

The letters *H* and *L* which stand for centers of high and low atmospheric pressure are scattered unevenly over the maps. They are called ***semi-permanent high- and low-pressure cells,*** or, more briefly, semi-permanent highs or semi-permanent lows, and are caused by unequal heating of the earth's surface. They are called semi-permanent since they remain more or less in one general locality throughout a season. The small movement they do have, however, is very important in connection with long-range weather forecasting, discussed in Chapter 14. These cells are also important because they determine wind direction and velocity. All winds and convection currents in the atmosphere are the result of differences in atmospheric pressure. Like water running down hill, ***the air in the atmosphere flows from areas of high pressure to areas of low pressure,*** and as one would expect, ***the greater the difference in atmospheric pressure between adjacent highs and lows, the stronger is the wind.***

If the earth did not rotate, the winds would blow directly ***away*** from a center of high pressure and directly ***toward*** a

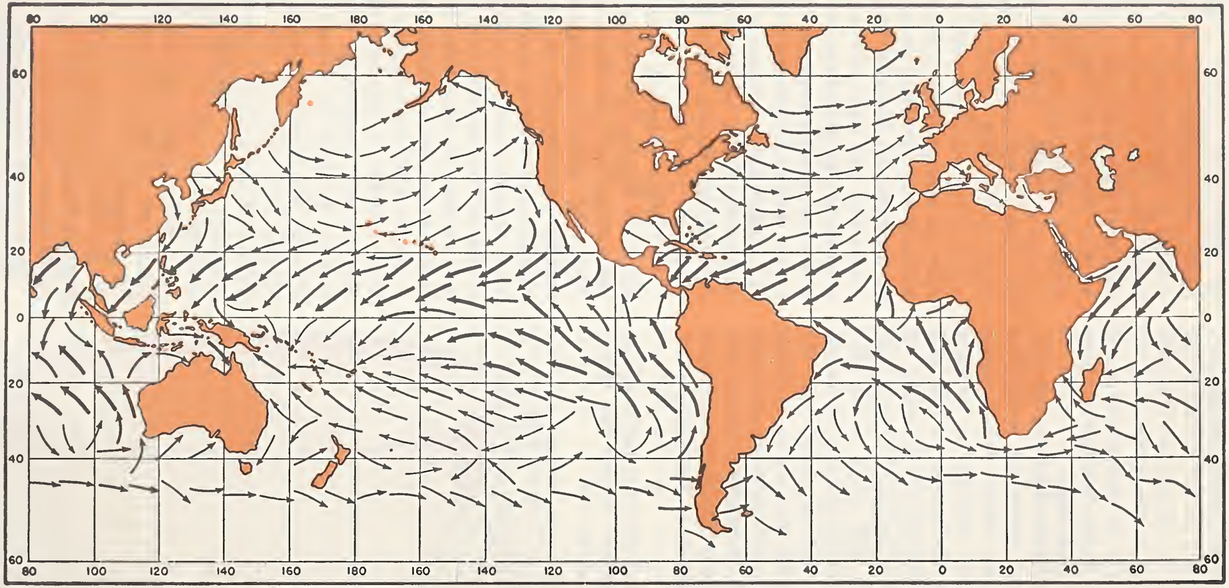


Fig. 11-6A. Prevailing winds over the oceans in winter.

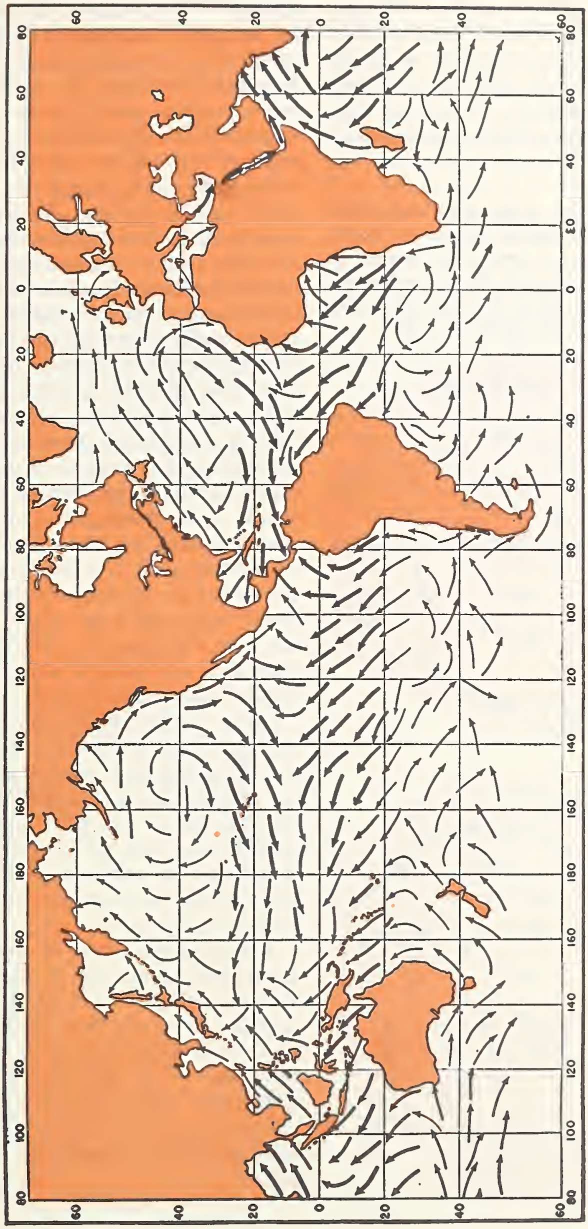
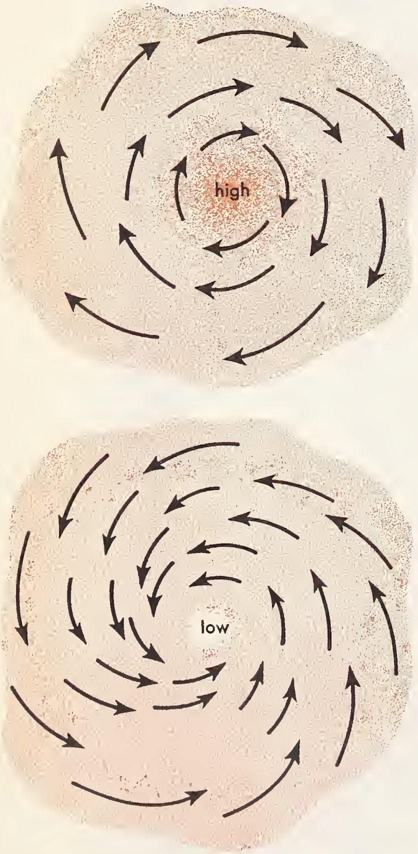


Fig. 11-6B. Prevailing winds over the oceans in summer.

center of low pressure. However, as a result of the earth's rotation, the air spirals clockwise outward from a high, and counterclockwise into a low as shown in Fig. 11-7. At least this is true in the Northern Hemisphere. What of the Southern Hemisphere? Here the direction of the spiral in both highs and lows is reversed. Why is this so? Let us examine, for example, the apparent rotation of the earth when viewed first from the South Pole and then from the North Pole. In looking at the earth's surface from the Southern Hemisphere one is, in a sense, looking "up" from "below" at the rotating earth and it appears to rotate clockwise whereas in the Northern Hemisphere one is "looking down from above" and the earth's rotation appears to be counterclockwise. In other words, all rotations appear to be reversed when one moves from one hemisphere to the other. Perhaps you will understand this apparent reversal in rotation a little more clearly if you consider a wheel spinning horizontally in a clockwise direction when you are looking down on it. If you now get under the wheel and look up at it, it will appear to be rotating in a counterclockwise direction. Try this experiment with some wheel such as a bicycle wheel. The arrows in Fig. 11-7 show the spiral motion about highs and lows. Look carefully at the arrows showing the direction of flow near them and you will observe the spiral motion which they indicate.

Fig. 11-7. Looking down at the earth's surface over a center of high atmospheric pressure (a "high") [the top diagram] and a center of low pressure (a "low") [the bottom diagram]. The air spirals outward, away from the center of a high, but spirals inward toward the center of a low.



Other smaller highs and lows move rapidly over great distances. These appear and disappear in a matter of days and are responsible for much of our day-to-day changes in weather. They are considered more fully in later chapters and should not be confused with the large, slow-moving, semi-permanent highs and lows described above.

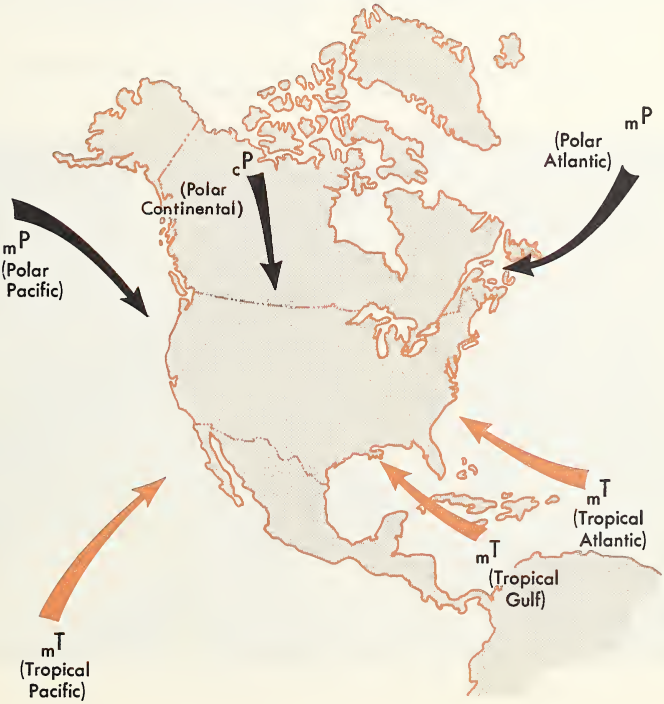
A close examination of weather maps shows that many high pressure cells lie close to either 30° S. latitude or 30° N. latitude. It is the flow of air spiraling away from these highs that give rise to the trade winds and stormy westerlies. The general position of the cells is pretty much determined by the position of large land masses and oceans on the earth's surface.

AIR MASSES

The flow of warm air from the Equator and its return as cold air from the polar regions is, as you can see, a highly complex system. Moreover, the flow is not steady since it is interrupted by the movement of ***air masses.*** At irregular intervals large masses of air, perhaps 500 to 5000 miles in diameter and thousands of feet deep, move southward from the arctic or northward from near the Equator. These air masses, as they are called, are formed chiefly near high-pressure cells in the polar regions and near the Equator. High-pressure cells are areas relatively free from winds where the air remains stagnant long enough to acquire temperatures and humidity characteristic of the surface over which it lies. In this way an air mass acquires a weather personality different from that of other air masses lying over other surfaces. For example, an air mass that forms over northern Canada will be cool, because of its northerly birthplace, and dry because it is formed over land where there is little open water from which evaporation can take place. Even in winter, when the ground is covered with snow, this air will be dry. Why? On the other hand, an air mass that originates over tropical seas will be warm and moist and so it may produce heavy rains.

Different air masses do not, as a rule, mix; they remain separate bodies of air as they move south or north. But the boundaries between them often become areas of precipitation and sometimes give

Fig. 11-8. The sources of air masses that move across North America.



rise to large storms. However, as air masses move they pick up some of the characteristics of the surface over which they travel. A cold air mass, moving south from northern Canada, becomes drier and warmer in its lower part near the surface of the earth as it moves over the land of more southerly latitudes. Why does it become drier? This is because as it moves south it gets warmer and therefore its relative humidity falls. On the other hand an air mass moving north from the Gulf of Mexico becomes cooler. Consequently the relative humidity rises, giving rise to clouds and precipitation, and muggy weather. Obviously, the faster an air mass moves, the less time it has to be modified by the surface over which it flows and the more it retains the characteristics of its place of origin.

Air masses are named according to their place of origin and are indicated on weather maps by symbols. The air masses having the greatest effect on the weather of North America are listed below with their symbols and most important characteristics.

Polar Continental
cP cold, dry, clear

Polar Maritime
mP cool, moist, hazy

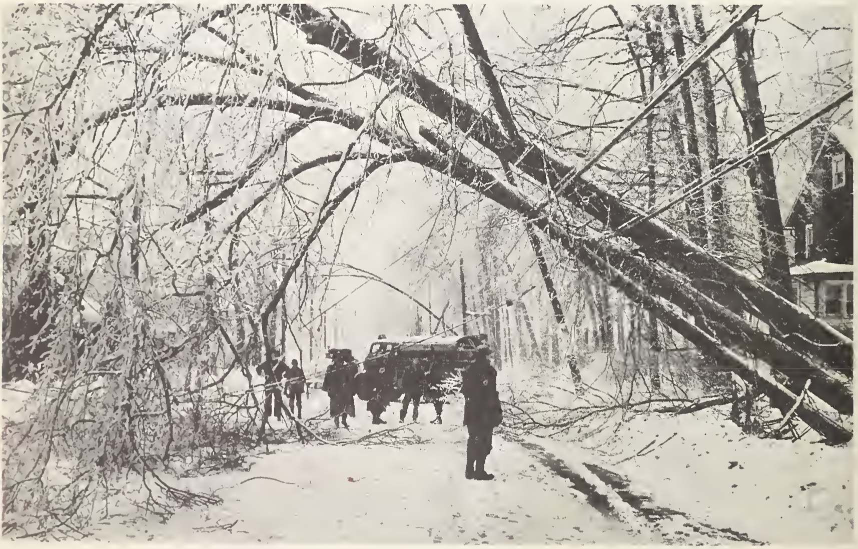
Tropical Maritime
mT warm, moist, hazy

The sources of different air masses throughout North America are shown in Fig. 11-8.

Polar Continental air from northern Canada is responsible in mid-winter for cold waves that often cause temperatures below freezing and clear, cloudless skies as far south as Texas and Florida. In summer, Polar air masses bring crystal-clear cool days to much of the North American continent.

Most of the warm air masses that move over the prairies and eastern Canada come

Fig. 11-9. A sleet storm often disrupts communication. Sleet results when rain falls through a layer of the atmosphere that is below freezing.



from the tropical regions in the Atlantic and the Gulf of Mexico. These outbreaks of Tropical Gulf or Tropical Atlantic air have absorbed large quantities of water vapor from the ocean. In summer this air brings hot, uncomfortable weather and rain. In winter it is responsible for most of our snowfall and the occasional midwinter thaw and rain, particularly in eastern Canada.

In the prairies the annual precipitation is light, but fortunately most of it falls during the important growing season. In Manitoba rain is chiefly due to the northward movement of moist tropical air. Farther west in Saskatchewan and Alberta rain and thunderstorms are caused by cool Arctic air lifting the warm air over the prairies. When these cool waves are infrequent there is less rainfall, and farmers suffer because their crops are below normal.

The West Coast of Canada and the United States has relatively cool summers and warm winters because it is visited mostly by Polar Maritime air masses. These air masses, like the ocean where they originate, never get very hot or very cold. The moisture from these masses is deposited over the western half of British Columbia, giving abundant rain. These Pacific air masses result in the "chinook" winds that cause the rapid changes in the weather of Alberta. The air, dry after it crosses the mountains, comes down to the lower levels of the prairies. As it does so, the air pressure is increased from the low pressure above the mountains to the higher pressure of the prairies. This compression heats the air (you are familiar with this increase in temperature when you pump up a tire), so that it is at a higher temperature than when it left the Pacific coast. Also, on the east side of the mountains, it replaces cold Arctic air that had previously lain there. This change in air mass causes a sudden rapid rise in temperature. In extreme cases, the temperature can rise 40°F in five minutes.

Things to Remember

The ***direction of winds*** is determined by unequal heating of the earth's surface and the rotation of the earth.

Trade winds blow steadily from the northeast in the Northern Hemisphere and from the southeast in the Southern Hemisphere. Trade winds are found between latitudes 10° N. and 25° N. and 10° S. and 25° S.

The ***doldrums*** are a belt around the Equator where there is little or no wind.

Winds blow from areas of high pressure to areas of low pressure.

The ***rotation of the earth*** causes winds to veer away from the straight line connecting points of high and low pressure.

The greater the ***difference in pressure*** between adjacent highs and lows, the stronger the wind.

A ***cold air mass*** is a large cold dry mass of air that originates in the arctic or antarctic regions.

A ***warm air mass*** is a large, warm and usually moist mass of air that originates in the semitropics.

Questions

GROUP A

1. Name three factors that help determine the weather over the world.
2. What causes wind?
3. What and where are (*a*) the trade winds, (*b*) the doldrums?
4. What is (*a*) a prevailing wind, (*b*) a monsoon?
5. What are air masses?

GROUP B

6. Why does it take less time to fly eastward across the country than it does to fly westward?
7. Explain why the trade winds blow from the northeast in the Northern Hemisphere.
8. What happens if you walk from the center to the edge of a rotating merry-go-round? What does this effect have to do with the direction of winds on the earth's surface?
9. What are (*a*) the horse latitudes, (*b*) the polar easterlies, (*c*) the stormy westerlies?
10. Do you think other planets would have the same kind of winds as earth? Explain.
11. What is the effect of the rotation of the earth on wind direction?
12. Name and give the symbol and describe three important kinds of air mass.
13. What are semi-permanent pressure cells?
14. Why is the west coast of Canada cool in summer and warm in winter?

Things to Do

Determining the direction of a prevailing wind · Locate a tall building with a flag that is flown every day. From the direction in which the flag blows, determine and record the direction of the wind twice a day for two weeks. Make your observations at the same time every day. From your records, determine the prevailing wind direction during this period.

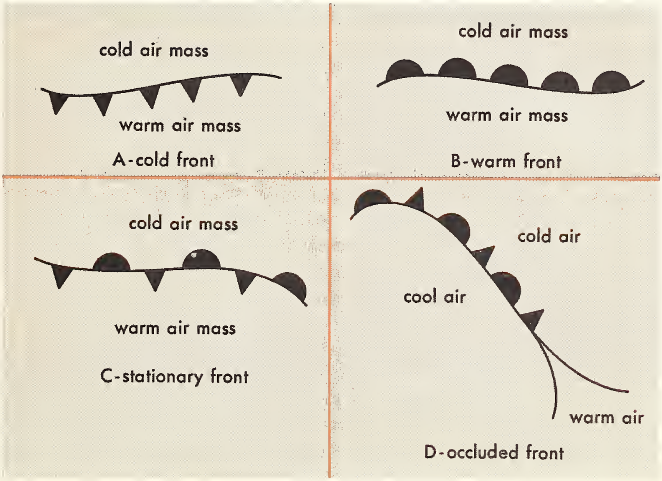
12

FRONTS AND STORMS

IN THE TEMPERATE ZONE the weather changes frequently, and it is not the same everywhere at the same time. These changes are due principally to the movements of air masses. Rapid changes in weather and most storms occur when the boundary between a cold air mass and a warm air mass passes overhead. This boundary is called a ***front.***

Suppose two adjacent large air masses of different temperature are moving across the country. If the warm air mass pushes the cold air mass before it, the moving boundary between the air masses is called a ***warm front.*** If the colder air pushes back the warmer air, a ***cold front*** is formed at the moving boundary. If two adjacent air masses are nearly sta-

Fig. 12-1. Symbols used on weather maps to indicate the four kinds of fronts. The points and semicircles point in the direction that the front is moving across the earth's surface. In C, a stationary front, the fact that they point both ways indicates that the front is not moving.



tionary, then neither air mass pushes back the other, and the boundary cannot be either a cold or a warm front. Instead it is simply called a ***stationary front***. There is still a fourth type called an ***occluded*** front which is associated with storms and is described on page 116. The symbols used to indicate each of the four types of fronts on weather maps are shown in Fig. 12-1.

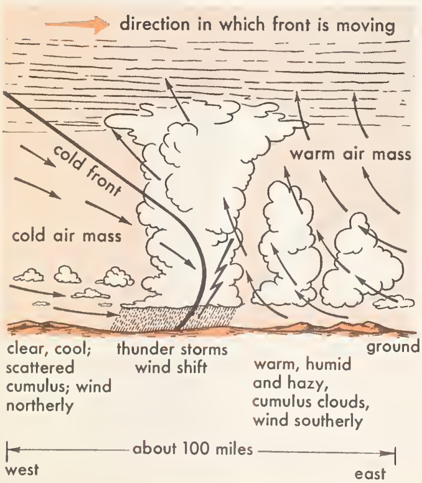
A Cold Front. A vertical cross section of a cold front is shown in Fig. 12-2. At the left is the mass of cold, clear air which is moving over the land towards the right, or eastward. In moving it wedges itself under a warm, humid air mass to the east.

Notice the peculiar bulge in the front near the ground. How do you account for it? The friction between the moving air near the ground and the ground itself slows down the air near the surface so that cold air above pushes ahead, overrunning the front near the ground. In this way the bulge is formed. A slowly moving cold front does not form a bulge, since at low speeds friction with the ground is negligible.

When the moving cold air pushes under warm air, the warm air, being lighter, is forced up. In rising, the warm air expands because of the lower atmospheric pressure at higher altitudes, and in expanding, the air cools. As it does so, water vapor begins to condense into billions of tiny drops to form ***cumulus*** clouds. The word ***cumulus*** means a heap and well describes this type of cloud which is big and fluffy, and flat on the bottom. There are several types of cumulus clouds, but they all resemble to some degree those shown in Fig. 12-3. If the warm air mass contains enough moisture, the cumulus clouds may tower into huge "mountains" of condensed vapor, called ***cumulonimbus*** clouds (see Fig. 12-18) and develop into thunderstorms. (***Nimbus*** is the Latin word for rain cloud.)

Let us suppose that you happen to live at a point located somewhere to the east of an approaching cold front, that is, on the right-hand side of the front shown in Fig. 12-2. What changes in the weather would you notice as the cold front approaches? Since you are in the warm air mass, the weather around you is likely to be warm and humid, the air is hazy and the wind is probably southerly. If you look toward the west, you will see cumulus clouds which become larger and thicker until a solid bank of

Fig. 12-2. A vertical cross section of a cold front.


direction in which front is moving
cold front
warm air mass
cold air mass
ground
clear, cool; scattered cumulus; wind northerly
thunder storms wind shift
warm, humid and hazy, cumulus clouds, wind southerly
about 100 miles
west
east

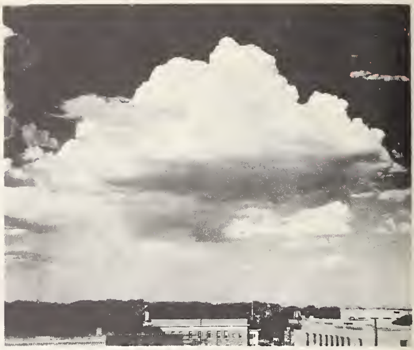


Fig. 12-3. Cumulus clouds are formed by the condensation of water vapor in rising warm air. They are flat on the bottom (the condensation level of the rising air) and puffy on top.

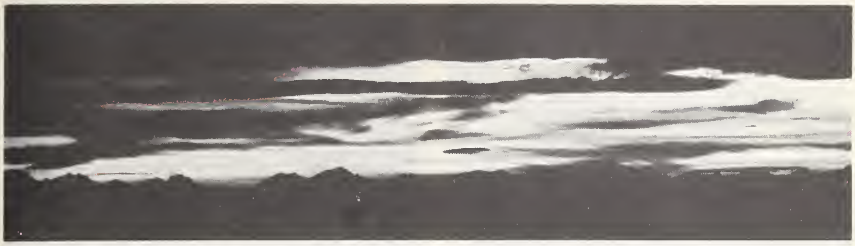


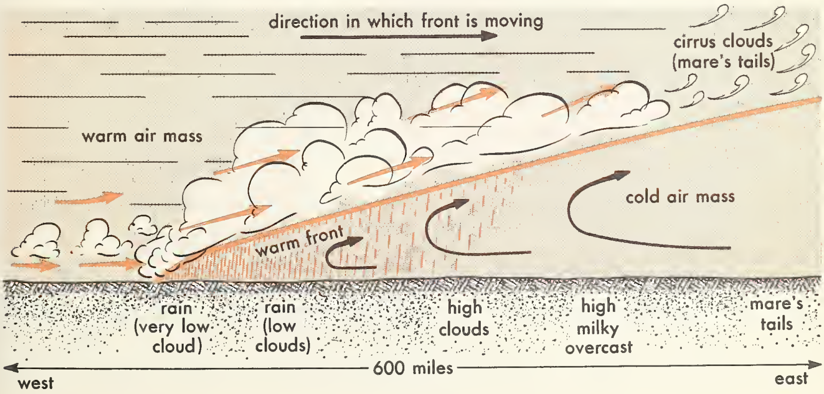
Fig. 12-4. A chinook arch in southwestern Alberta. The appearance of such a cloud formation in the southwest part of the sky means that milder weather is on its way.

dark cloud comes over the western horizon. When the bank of cloud arrives nearly overhead, the wind veers from the south to the northwest. Why is this? This northwest wind may blow in sharp gusts which may be accompanied by violent but brief rain squalls. After the cold front has passed overhead, the rains stops, the wind settles into a steady northwest breeze, and the clouds begin to thin out. The air that now surrounds you is that of the cold air mass. In summer, it is cool and clear and brings relief from the hot, sticky air that probably preceded it. In winter it usually marks the beginning of a series of cold, clear days. The time that elapses, from the first appearance of the cold front in the west to its disappearance over the eastern horizon, may be only an hour or so.

Cold fronts, like people, vary a great deal in their "personalities." Some move rapidly with great violence. Some may be accompanied by thunderstorms, hail, blinding snow squalls, violent winds and even tornadoes; others are so mild that only a meteorologist is aware of their passage. If the warm air mass contains little moisture, there may be no rain and little cloudiness along the cold front. Moreover, if the two air masses are of nearly the same temperature, there will be little or no temperature change as the front passes.

A Warm Front. The changes that occur along a warm front are more gradual, less violent, and last longer than those along a cold front. The first sign of the approach of a warm front is the appearance of ***cirrus*** clouds high in the sky. (The word

Fig. 12-5. A vertical cross section of a warm front.



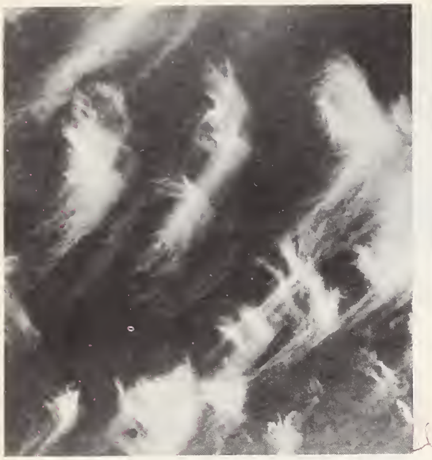


Fig. 12-6. Mare's tail—a form of cirrus cloud that often gives warning of an approaching warm front. Cirrus clouds occur at very high altitudes and are composed of tiny ice crystals.

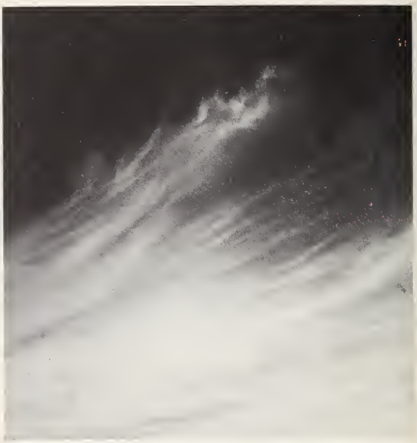


Fig. 12-7. As the warm front approaches, the mare's tails fuse into a sheet of cirrus cloud called ***cirrostratus.***

cirrus means a curl.) These clouds first appear as thin, semitransparent wisps called ***mare's tails*** and are shown in Fig. 12-6. They form over a warm front at high altitudes where the temperature is well below freezing and are therefore composed of tiny ice crystals. Cirrus clouds are blown far ahead of a warm front, since winds at high altitudes are not impeded by ground friction and are therefore stronger. These clouds often give one or two days warning of the approaching front.

After the appearance of mare's tails, the sky slowly becomes covered with a thin layer of cirrus clouds, often appearing as no more than a whitish haze in the sky through which the sun can shine.

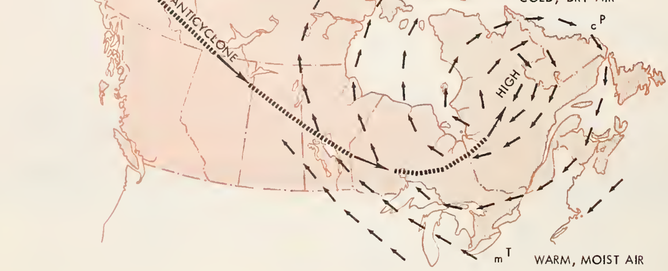


Fig. 12-8. The pattern of airflow about a high, or anticyclone, centered in northern Quebec.

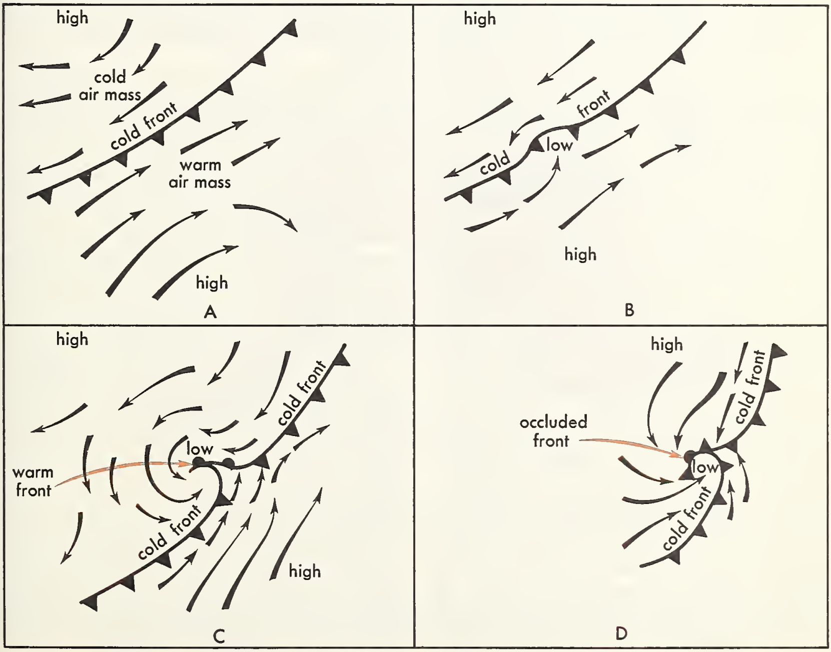
These are called ***cirrostratus*** clouds as shown in Fig. 12-7. (The word ***stratus*** means a spreading out.) The clouds slowly become thicker and lower until, finally, the sky is darkened by a heavy layer of ***nimbostratus*** clouds, and steady rain begins to fall. The rain increases in intensity until the front actually arrives and the warm air mass makes its appearance. Then the rain gradually stops. The clouds thin out until only a few puffy cumulus clouds remain, and we find ourselves in the middle of a mass of warm, humid, and hazy air typical of a warm air mass.

The passage of a warm front may take one or two days, and the total amount of precipitation (rain or snow) is much greater than the brief shower that accompanies a cold front. One reason for this is that warm fronts generally move across the land more slowly than cold fronts. Another reason can be understood by comparing the lines marking the cross sections of the fronts in Figs. 12-2 and 12-5.

You will notice that, as a result of slower speed and less friction, the slope of the warm front is not nearly as steep as that of the cold front. And since most rains occur near the front, a much larger area of precipitation is covered by the warm front than by the cold front. In winter the passage of cold and warm fronts is much the same as in summer, except that we may have snow instead of rain.

Cyclones and Anticyclones. Let us now consider the effect of atmospheric pressure on weather. Winds, as we have seen in Chapter 11, are caused by a difference in pressure, and the prevailing winds are due to the semi-permanent high- and low-pressure cells. But the changing winds which accompany the day-to-day weather changes, so common in the

Fig. 12-9. The development of a cyclone



temperate zones, are the result of smaller, rapidly moving lows and highs called ***cyclones*** and ***anticyclones*** respectively.

Anticyclones generally bring fair weather. Can you tell why? Fig. 12-8 shows an anticyclone moving from west to east across Canada. Notice how it first brings cold, clear weather with northerly winds (cold, dry, cP air sweeping down from the north, clockwise about the center of the high) which gradually changes to warm, moist air with southerly winds (mT air sweeping up from the south, clockwise, about the center of the high).

Cyclones, on the other hand, frequently bring stormy weather because they are usually accompanied by rapidly moving fronts. As these lows approach, the barometer begins to fall, and bad weather can be expected. Although severe storms often accompany a cyclone, many cyclones pass by with no more than light winds and occasional brief showers. Cyclones and anticyclones are extensive in area. A cyclone may be several hundred miles in diameter; the anticyclone is frequently much larger.

Cyclones that cause our stormy days originate along a cold front. Fig. 12-9 shows how a cold front, as it appears on successive weather maps, may change and develop into a cyclonic storm. In Fig. 12-9A the cold front moves slowly toward the southeast across the map before there is any indication of a developing storm. In Fig. 12-9B, some hours later, the cold front has developed a slight bulge, accompanied by a slight drop in pressure near the center of the bulge. As the bulge grows larger, warm air moves into it and cold air moves in a curve around it, the pressure in the center dropping still more until finally in Fig. 12-9C we have a fully developed cyclonic storm with strong winds, low pressure and considerable precipitation along the fronts. You will notice that, as the warm air pushes its way into the bulge, it begins to push part of the cold front back thereby changing this section of the cold front into a warm front.

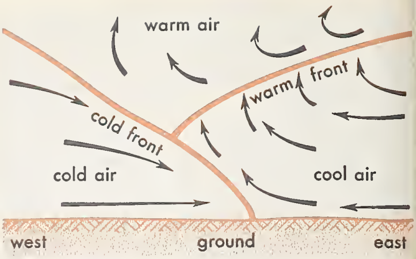


Fig. 12-10. A vertical cross section of an occluded front. The warm front has been lifted off the ground near the center of a cyclone by the rapidly moving cold front.

Eventually, since the cold front near the center of the low moves faster than the warm front, it overtakes the warm front, pushes under it and lifts it off the ground giving rise to an ***occluded front*** as shown in Fig. 12-9D. (A vertical cross section of an occluded front is shown in Fig. 12-10.) Once the occluded front starts to form, the storm begins to dissipate and die as the low fills up with air. The pressure rises, the wind drops, and finally the rain ceases and the storm disappears altogether.

The development of a full-fledged cyclone takes from one to three days and it may last for several days or a week or more before it becomes occluded and finally disappears. Cases have been recorded where cyclonic storms of this type lasted long enough to be followed on weather maps over a path that took them completely around the world back to their point of origin before dying. This however is a rare occurrence.

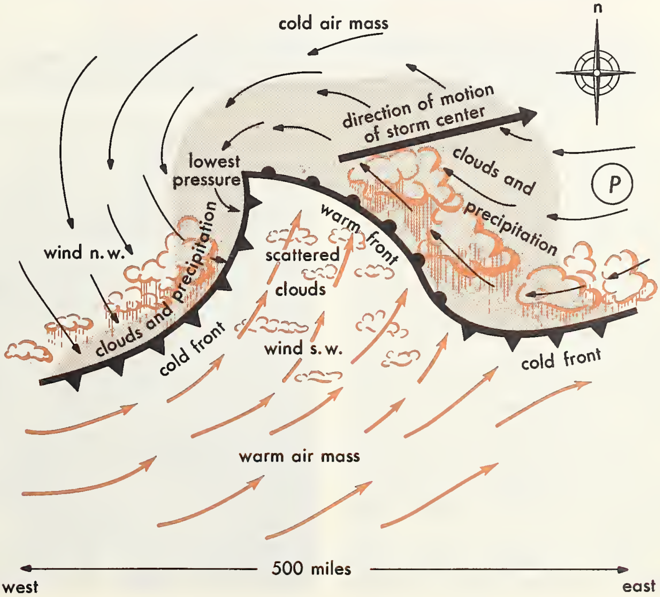
What changes in weather might we expect as a fully developed cyclone approaches and passes over us? Let us assume that the cyclone is moving eastward and that its center passes just to the north of us; in other words, we are located at point *P* on the map of the storm shown in Fig. 12-11 so that the section of the storm along the dotted line will pass over us. First we would observe the wind direction to be southeast. The barometer would fall as the center of low pressure approached, and the wind would rise. Why would the wind rise? The clouds would thicken and it would begin to rain as the warm front approached. Can you explain why? Then, after the warm front had passed we would have warm, unsettled and partially clear weather for perhaps half a day. Then, after a brief but heavy shower accompanying the cold front, the barometer would begin to rise, the wind would swing around to the northwest, the temperature would fall, the air would become clear and the good weather of an anticyclone would probably follow. Can you give the reasons for these observations?

The cold fronts along which cyclonic storms originate occur mostly in the temperate zones between large polar and tropical air masses. It is for this reason the climate of the temperate zones is much stormier and more variable than the climate of the tropics. This also explains why the prevailing westerly winds of the temperate zones are known as the ***stormy*** westerlies.

Hurricanes. During the late summer, people in the coastal regions from Florida to Nova Scotia and Newfoundland are apprehensive. This is the time for hurricanes, or typhoons as they are called in Asia. In some seasons, seven or eight will approach the coast and two or three may strike the mainland causing great damage. What are hurricanes? How do they get started? Can anything be done to deflect or control them?

Hurricanes are similar to cyclones in effect. But unlike cyclones, they always originate over water near the Equator, and never over land, nor do they form along a cold front. Scientists are not exactly sure what starts them. One cause of their violence is undoubtedly the strong heating of the ocean near the Equator that results in a spiraling updraft covering an area of hundreds of square miles. The winds of a hurricane are much more violent than those of a

Fig. 12-11. A cyclonic storm as it appears on a weather map.



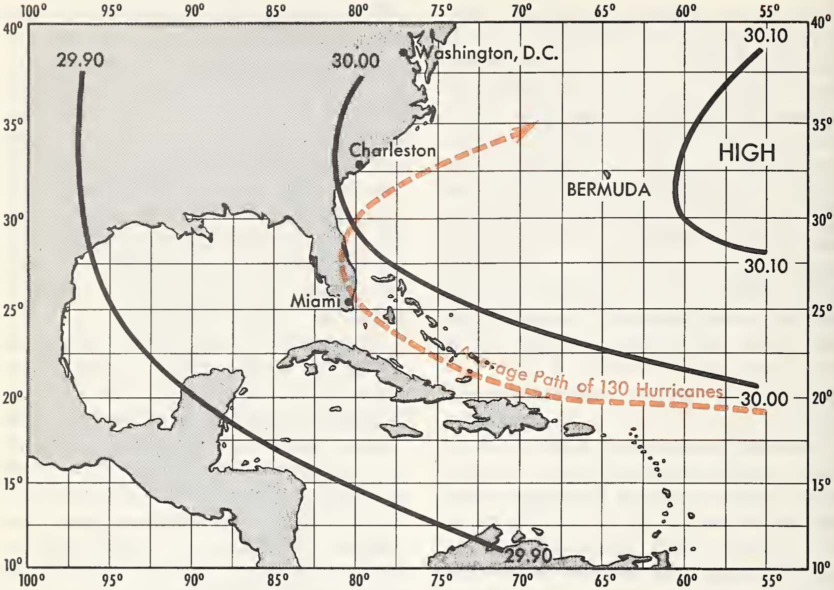


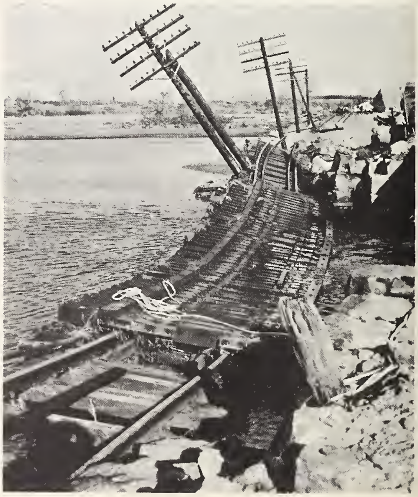
Fig. 12-12. The average path of 130 hurricanes (dashed line). The solid lines show the average atmospheric pressure over the area of the map. The high near Bermuda is the semi-permanent high known as the "Bermuda High." Notice how hurricanes tend to follow a path that takes them around this high.

common cyclone. They have been known to reach a velocity of 130 miles per hour. Of course this refers to the speed of the spiraling winds. The hurricane itself moves along its path at a slow speed of 10 to 50 miles per hour.

Hurricanes rarely reach very far inland. This is because the friction offered by rough and tree-covered land is so great that it rapidly uses up the energy of the hurricane and reduces its wind velocity until it wears itself out. Moreover, hurricanes are dependent upon water for their energy; they gain energy from the latent heat of the condensation of evaporated ocean water. As the hurricane moves over land this supply of energy is cut off. In 1954, however, one hurricane reached Toronto, bringing torrential rains, high winds, and floods.

Florida and the other southeastern coastal states have long been the targets

Fig. 12-13. The New England hurricane of 1938 did this damage to a railroad track.



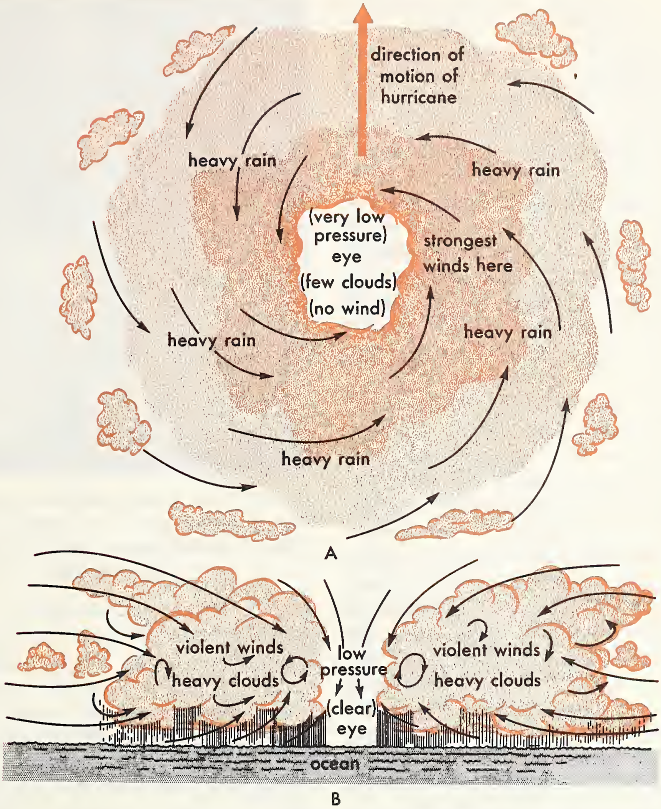


Fig. 12-14. The structure of a hurricane; looking down from above the center (A) and looking at a vertical cross section (B).

of hurricanes but, in recent years, quite a few of these storms have by-passed this area and struck the middle-eastern and northeastern Atlantic states. The paths of some of these hurricanes in recent years is shown in Fig. 12-12. You will notice that the average path sweeps up the coast in a large arc into the North Atlantic, skirting around the edge of the Bermuda High, a semi-permanent high-pressure cell, usually centered near Bermuda. The path of any given hurricane is governed, not only by the Bermuda High, but also any migrating highs and lows that happen to be near when the storm moves north. The storm is pushed aside by high-pressure systems and attracted into low-pressure areas. The United States Weather Bureau maintains an elaborate hurricane advisory and warning system that follows the storms soon after they are born, east of the West Indies, and, from a prediction of the movements of other lows and highs, broadcasts warnings to shipping and coastal areas concerning the probable course of the hurricane.

A diagram of the structure of a hurricane is shown in Fig. 12-14. The center of the storm, or **eye** as it is called, is an area, perhaps 40 miles across, with clear skies and little or no wind. Surrounding the eye on all sides, are great towering ramparts of swiftly moving black clouds. The strongest winds are just beyond the eye, as is also the heaviest rainfall. Since the storm has a forward motion in addition to the spiral motion of air around its center, the eastern side contains the strongest winds. Why is this? It is because, on this side, the forward speed is added to the speed of the spiraling air while on the western side it is subtracted from the circular speed. Thus, for example, hurricane Carol (hurricanes are usually given girls' names by the U.S. Weather Bureau) of 1954 had winds spiraling around the center at about 90 miles per hour and an unusually rapid forward motion up the east coast of 40 miles per hour. The maximum wind on the east side was therefore about 130 miles per hour while the highest on the west was about 50 miles per hour.

What happens weatherwise, as a hurricane approaches and its center passes directly over us? The first indication of the approaching storm, if one is on the coast, is the gradual building up of very heavy seas a day or two before the actual arrival of the storm. Large waves, caused by the hurricane's winds, race a thousand miles or so ahead of the storm. Then there is a period of light wind from the east, a drizzle, and perhaps fog as the barometer slowly falls. Later the rain and wind increase and the barometer falls rapidly until the hurricane strikes with full force, bringing torrential rain and gusty east winds in excess of 70 miles per hour. Then, very rapidly, the wind dies, the sky clears and the storm appears to be over. But a look at the barometer shows it to be very low and not yet rising, which indicates that this is the eye, the center of the storm. Very quickly the storm begins to rage again with violent winds now blowing from the west. How can you explain this? The barometer now begins to rise, the winds abate and the storm finally passes leaving warm dry air in its wake.

The total energy expended by a hurricane is almost unbelievable. It is hundreds of times that of a hydrogen bomb. A hydrogen bomb, however, releases all its energy in a fraction of a second and causes total destruction over a small area. A hurricane, on the other hand, lasts for over a week and spreads its violence throughout an area several hundred miles wide and several thousand miles long, causing widespread but not total destruction.

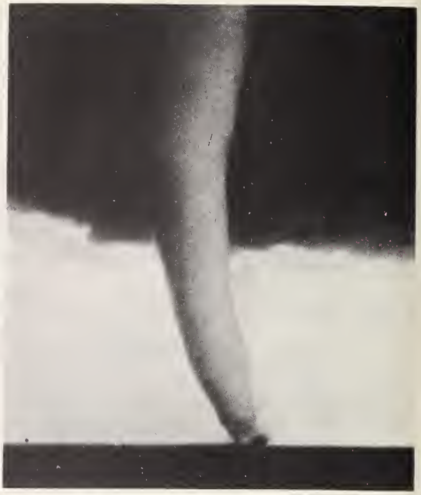


Fig. 12-15. The whirling funnel-like cloud of a tornado in Kansas. The air is spiralling around the center of the long cloud at speeds of several hundred miles per hour.

Tornadoes. Sometimes on a hot day in the midwestern United States a funnel-shaped cloud may be seen. Usually it seems to be suspended from a dark storm cloud and extends downward to the earth. This is an ominous sign. Within this funnel the air spirals upward in violent motion. It is a ***tornado,*** the most violent of all storms. A tornado formed out at sea is called a waterspout.

Tornadoes are small local cyclones that start on very hot days, near cold fronts and are generally accompanied by thunderstorms. They are sometimes only a few hundred yards in diameter and generally travel only a few score miles. But

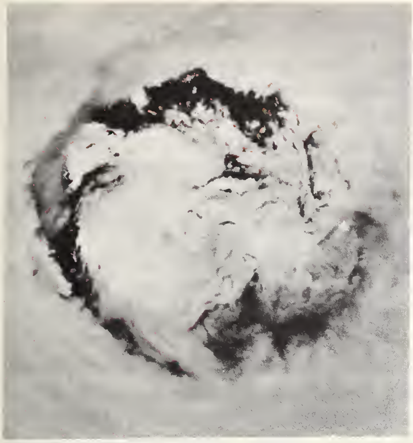
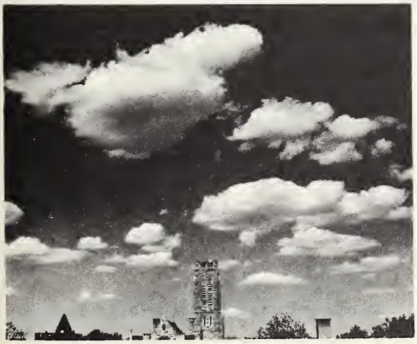


Fig. 12-16. The eye of a hurricane. The dark areas are the ocean showing through the partly cloudy area of the eye.

they destroy nearly everything in their path. In the winds, which may have a velocity of well over 200 miles per hour, houses may be smashed, steel girders twisted, and freight cars blown from their tracks. Small ponds have even had all the water sucked out of them by a tornado. In such a case the fish, frogs and debris from the pond are also sucked up. They fall to the ground some distance away, so that it literally rains fish and frogs!

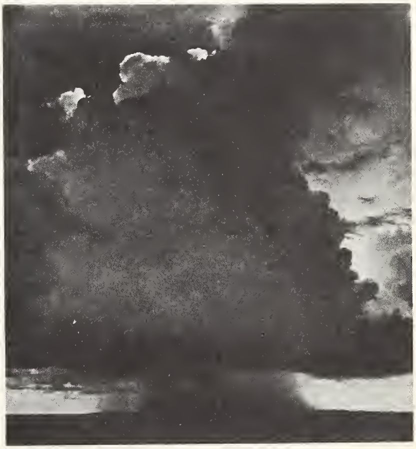
Fig. 12-17. Small cumulus clouds that do not grow rapidly or very large are called fair-weather cumulus clouds.



Thunderstorms. Hot summer days as well as cold fronts often give rise to thunderstorms. On such days, the air just above the ground becomes very warm, expands, and rises, forming a fountain of warm air rising more than a mile into the atmosphere. To replace the rising air, cool air rushes in along the ground causing strong, gusty squalls of wind. As the warm air rises and expands it becomes cooler. If the air cools to saturation, condensation begins, and cumulus clouds are formed. Sometimes they look like Fig. 12-17, which are called ***fair-weather cumulus*** because they do not grow very large and appear on warm sunny days. If, however, there is much moisture in the rising air, a great deal of latent heat is released when condensation takes place. This heats the air, making it still lighter, so that it rises still higher and forms large, towering ***cumulonimbus clouds*** like those in Fig. 12-18. These produce heavy, though usually brief, showers accompanied by thunder and lightning. Such a towering mountain of condensed water vapor is the heart of a thunderstorm.

Whenever possible airplane pilots try

Fig. 12-18. When cumulus clouds grow to heights of several thousand feet and become thunderstorms, they are called *cumulonimbus clouds.*



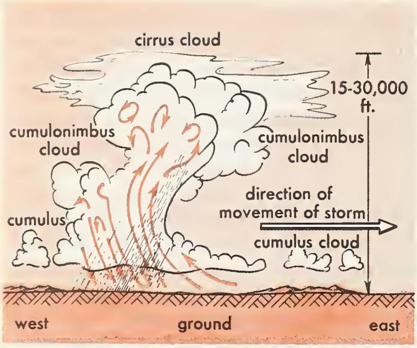


Fig. 12-19. Vertical cross section of a thunderstorm.

to avoid cumulonimbus clouds by flying around them because they are dangerous. The strong updrafts and downdrafts of air within these clouds may lift a plane suddenly upward and then drop it downward, sometimes with enough force to tear off a wing.

Hail. Under the right conditions, rain may be changed to ice and a hailstorm is formed. Hailstorms occur frequently; every year they are responsible for much damage to crops. Hailstones (or hail, as they are sometimes called) are formed in severe thunderstorms. Small raindrops are carried upward by strong updrafts. When they are carried above the freezing level some form into small spheres of ice. In the rising currents, these do not rise as rapidly as the surrounding air. Some of the moisture from the air then catches and collects on the hailstones, increasing their size. This continues until the hail becomes so large that it can no longer be supported by the updraft and so falls to the earth. With very strong updrafts, hailstones larger than hens' eggs can be formed. Some hailstones, if cut across, are found to be formed in concentric spheres. These arise if the stone in growing finds itself alternately in very moist and then in less moist parts of the cloud.

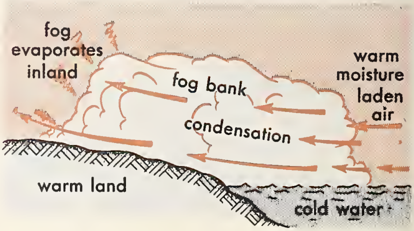
Although hail falls in most parts of southern Canada, the worst hail belt in the nation is found in Alberta in the area between Red Deer and Calgary. The pattern of frequency is very irregular in time and place. A number of sections along and just west of the railroad from Calgary to Red Deer average one hailstorm a year. The frequency decreases rapidly eastward so that along the Saskatchewan boundary a given area may expect no more than two hailstorms in a ten-year period. One particular thunderstorm cell will bring hail to a small strip possibly half a mile wide and 6 miles long. But a series of such cells will give hail along a strip one to two miles wide and 50 to 100 miles long during one afternoon.

water surface. When warm moist air from the Gulf Stream crosses the cold waters of the Labrador Current off Newfoundland, then Nova Scotia and the Grand Banks are usually enveloped in fog. In fact Cape Race is fogbound for as many as 158 days a year. Similar fogs occur along the coast of British Columbia. Dense, dirty and persistent fog occurs in Vancouver in November and December. On occasion a swirling fog almost paralyzes transportation facilities.

Fog. A fog is similar to a cloud, except that it is formed on the ground rather than a thousand or more feet up in the air. It is formed when warm moist air that is nearly saturated with water vapor is cooled by contact with a cold land or water surface.

Fogs make transportation by ship and by aircraft extremely dangerous. A great deal of money is spent each year on safety devices such as radar and weather forecasting in an effort to prevent possible accidents in thick fogs.

Fig. 12-20. How a coastal fog bank forms.



Things to Remember

A ***front*** is the boundary between two air masses. Rain often occurs along fronts.

Along a ***warm front*** the warm air mass pushes back the cold air mass.

Along a ***cold front*** the cold air mass pushes back the warm air mass.

A ***stationary front*** is the boundary between two stationary air masses.

An ***occluded front*** occurs when a cold front catches up with a warm front and pushes under it.

A ***cyclone*** is a large whirlpool of air with low pressure at the center and the air spiraling counterclockwise (in the Northern Hemisphere) toward the center. It has warm and cold fronts.

An ***anticyclone*** is a large whirlpool of air with high pressure at the center and the air spiraling clockwise (in the Northern Hemisphere) from the center.

Cirrus clouds are high, thin clouds consisting of ice crystals.

Cumulus clouds are large puffy clouds caused by rising air currents.

Thunderstorms are the result of strong vertical convection currents that form ***cumulonimbus*** clouds.

Hurricanes are cyclones that originate near the Equator. They are smaller than ordinary cyclones, have no fronts, but have much stronger winds.

Tornadoes are much smaller than hurricanes and much more violent. They form along strong cold fronts.

Hail occurs in thunderstorms from the freezing of raindrops.

Fog is a cloud touching the ground.

Questions

GROUP A

1. Describe the formation of cumulus clouds.
2. How may rain be produced?
3. What is (*a*) a cyclone, (*b*) an anticyclone?
4. What causes thunderstorms?
5. How is fog produced?
6. When the barometer falls, what kind of weather is likely to follow? What kind of weather is likely when it rises?

GROUP B

7. What are the principal differences between a cyclone, a hurricane, and a tornado?
8. Describe cumulus clouds and cirrus clouds.
9. Explain, with the aid of a diagram, how a thunderstorm is formed.
10. Describe, with the aid of diagrams, the life of a cyclone from its origin to its end.
11. What happens when a cold front approaches and passes over you?
12. What happens when a warm front passes over you?
13. What is an occluded front?
14. (*a*) Draw a diagram of the structure of a hurricane, (*b*) Where does a hurricane get its energy? Explain.

Things to Do

Recognizing different kinds of clouds · Examine the sky on two or three successive days and name the cloud formations that you observe. Are there any cumulus clouds? How were they formed? Do they remain in the same place? What finally happens to them? Try to answer the same questions for cirrus clouds. An excellent set of illustrations of many different kinds of clouds can be obtained free, by writing on school stationery, for the weather educational series of weather pamphlets, to the Director, Meteorological Branch, Department of Transport, 315 Bloor Street West, Toronto 5, Ontario. Your teacher should write for these.

13

WEATHER FORECASTING (1)

WEATHER INSTRUMENTS AND THE WEATHER MAP

In earlier chapters we have discussed rain, winds, fronts and storms, and their contribution in making up our daily weather. This information should make the daily weather changes fascinating to watch. Can you now recognize the conditions and say, "Here comes a cold front"? And can you then predict the changes that will occur as the front passes? Meteorologists, however, are interested in more than this. Their most important job is to predict the weather a day or two ahead. Will it rain within the next 24 hours? Can we expect strong winds to blow? Before a meteorologist can answer such questions he needs a good deal of information, most of which he gets from instruments.

Until fairly recent times the structure of storms was not understood and weather prediction was unreliable. A certain number of weather signs, such as a ring around the moon portending bad weather, had become familiar to men who sailed the seas or farmed the land —those most vitally concerned with weather at that time—but accurate weather forecasting had to wait upon an understanding of the causes of weather.

Benjamin Franklin was one of the first to become aware of an important aspect of the cyclonic nature of storms (a storm in which the winds spiral around the center of low pressure). It happened that a storm bringing northeasterly winds passed over Philadelphia, obscuring an eclipse of the moon that Franklin had wished to observe. Franklin naturally concluded that the storm had come from the direction of Boston. He was surprised to learn that the eclipse had been visible in Boston and the storm had arrived there ***after*** it had passed over Philadelphia. He also learned that the storm had arrived ***earlier*** at places to the south of Philadelphia. How could this be explained? He came to the conclusion that the only satisfactory explanation lay in assuming that the storm really moved from south to north as shown in Fig. 13-1, and did not come from the direction in which the winds were blowing. This was the first real understanding of the nature of cyclonic storms and a very important step toward more accurate forecasting.

The Meteorological Service. It must now be apparent that weather forecasting is a highly complicated procedure, because it depends on many factors. Can you re-

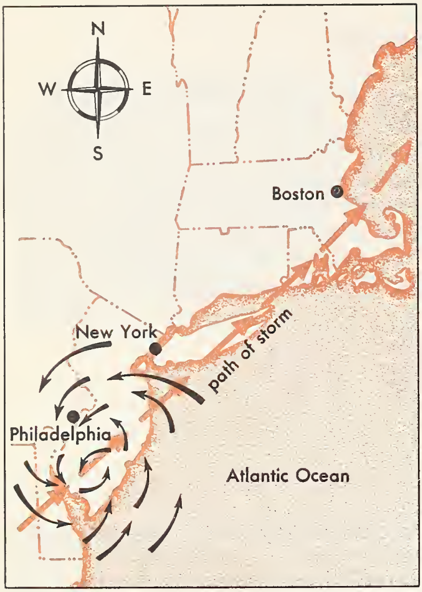


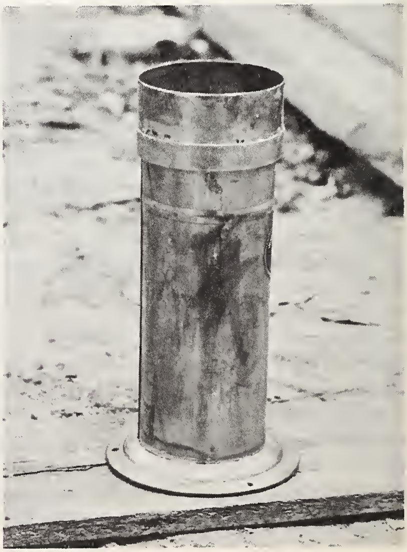
Fig. 13-1. Benjamin Franklin deduced that the winds blow in a spiral around the center of a storm from the fact that winds blew from the east in New York City while a storm moving northeast was approaching.

call some of the things a meteorologist must know in order to forecast the weather? First, he must have exact knowledge of the present weather surrounding the area. And, since air masses and fronts often move several hundred miles a day, he must know the position of all fronts and highs and lows, and how fast they are moving. He must also know the kinds of clouds present and the extent of cloudiness. In addition, he must also know the direction and velocity of the winds, temperatures, the amount and location of any precipitation, the amount of moisture in the air at different altitudes, and the barometric pressure at as many points as possible. With the help of these data he can predict with reasonable accuracy what changes in the weather will occur in the immediate future in a particular area.

The first service for gathering weather data was established in Europe during the last century. At that time the telegraph was not in widespread use. Warnings of approaching storms were relayed by signal flags from hill to hill across the countryside. The telegraph and, later, the teletype made possible the rapid collection of weather data and hence more accurate forecasting. Today, the forecaster receives his information from a weather map made up every few hours from reports that are teletyped to his office from nearly a thousand weather stations in Canada, the United States, Mexico and Cuba.

In every part of the world, official observers read their weather instruments simultaneously. These observations are made every six hours — at midnight, 6 A.M., noon, and 6 P.M., Greenwich Mean Time. In Labrador it may be a cold winter evening as the observer pulls on an over-

Fig. 13-2. A rain gauge consists of a narrow tube in which the rain is collected through a large funnel. The funnel forms the top part of the cylinder shown here; the tube is inside.



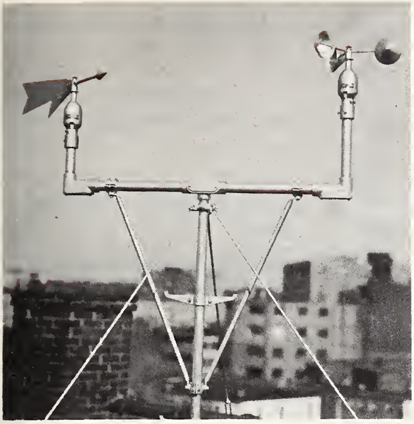


Fig. 13-3. A wind vane (left) and an anemometer (right). The three cups on the anemometer are blown around by the wind. Both wind direction and velocity are recorded automatically by means of electric circuits running from the instruments to an office in the building below.

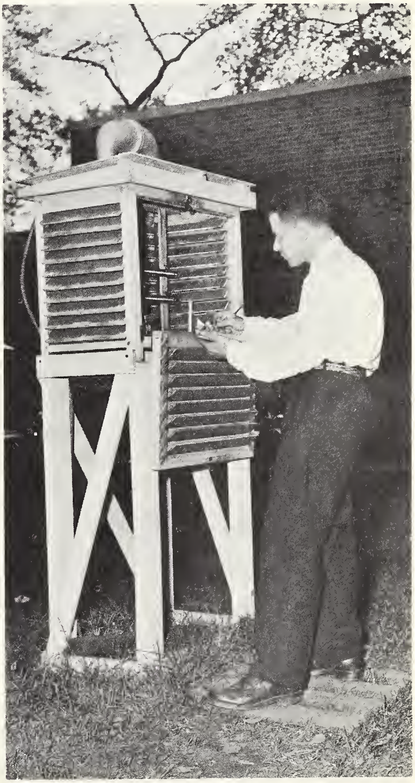
coat and trudges out to measure the snowfall; on Vancouver Island it may be a late sunny afternoon as an observer goes out at precisely the same time, in his shirt sleeves, to read his instruments. The observations are translated into an abbreviated code and sent by teletype to forecast offices across Canada and throughout the world. Additional reports are sent in at one-hour intervals from some airports to aid in preparing the weather map. Let us see how the most important weather information is observed and measured.

Precipitation. In meteorology precipitation means rain, snow or hail; in chemistry it means the formation of an insoluble substance from two soluble ones in solution; in both cases, precipitation is something that falls from above and settles on a surface. How is atmospheric precipitation measured? Rainfall can be measured by a ***rain gauge*** such as that shown in Fig. 13-2. It consists of a narrow tube in which the rain is collected through a large funnel. The tube and funnel are supported in a large outer can and the whole apparatus is placed in an open, level place. To measure the rainfall, the collected water is poured into a glass graduate. The scale on the glass graduate is adjusted to read directly the amount of rain in inches. With such a gauge, rainfall can be measured accurately to one-hundreth of an inch.

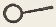
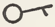
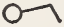
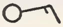
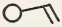
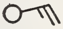
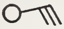
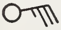
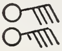
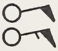
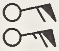
In winter the depth of freshly fallen snow is measured by averaging a number of readings of snow depth over a large area of level ground. A depth of ten inches of snow is reported as equivalent to one inch of rain.

Wind Direction and Velocity. Wind direction is indicated by a simple ***weather vane,*** a device which points in the direction from which the wind is blowing. The actual wind direction can be indi-

Fig. 13-4. Weather instruments installed at a weather station.



BEAUFORT SCALE OF WIND VELOCITIES

Beaufort Number	*Symbol*	*Type of Wind*	*Velocity in mph*	*Description*
0.	○	Calm	Less than 1	Smoke rises vertically.
1.		Light Breeze	1-2	Direction of wind shown by smoke drift but not by wind vanes.
2.		Slight Breeze	3-8	Wind felt on face; leaves rustle; ordinary vane moved by wind.
3.		Gentle Breeze	9-14	Leaves and small twigs in constant motion; wind extends light flag.
4.		Moderate Breeze	15-20	Raises dust and loose paper; small branches are moved.
5.		Fresh Breeze	21-25	Small trees in leaf begin to sway; crested wavelets form on inland waters.
6.		Strong Breeze	26-31	Large branches in motion; whistling heard in telegraph wires; umbrellas used with difficulty.
7.		High Wind (Mod. Gale)	32-37	Whole trees in motion; inconvenience felt in walking against wind.
8.		Gale (Fresh Gale)	38-43	Break twigs off trees; generally impedes progress.
9.		Strong Gale	44-48 49-54	Slight structural damage occurs (slates removed from roof).
10.		Whole Gale	55-60 61-66	Trees uprooted; considerable structural damage occurs.
11.		Storm	67-71 72-77	Rarely experienced; accompanied by widespread damage.
12.		Hurricane	Above 77	

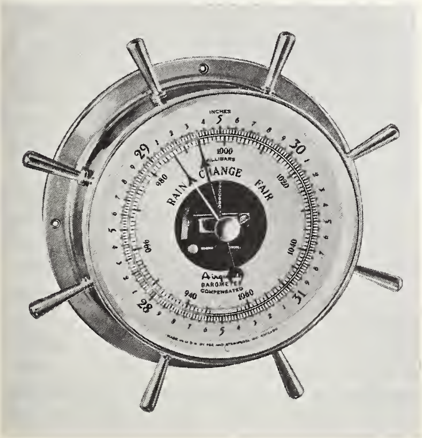
cated by electrical means on a compass dial located in some convenient place.

Wind velocity is measured by an ***anemometer,*** one type of which is shown in Fig. 13-3. The three cups rotate in the wind like a horizontal windmill, the speed of rotation increasing as the velocity of the wind increases. The speed of the shaft's rotation is recorded electrically on a meter dial that indicates the wind velocity in miles per hour.

Anemometers are standard equipment at every weather station, but the wind velocity can be estimated even when an anemometer is not available. This was first done by a sea captain named Beaufort, who devised a set of rules, now known as the Beaufort scale, for estimating the wind velocity at sea by the type and size of the waves. This scale has been adapted for use on land and is shown on page 128.

Barometric Pressure. The pressure of the atmosphere is measured most accurately by means of a mercury barometer, described in Chapter 3. A mercury barometer is part of the equipment in every weather station. But another type, called an ***aneroid barometer*** (see Fig. 13-5A and B) is often used because it is more convenient than a mercury barometer, though not as accurate. (The word ***aneroid*** means ***without liquid***.) It consists of a small disk-like metal box from which part of the air has been removed. If the atmospheric pressure increases, the top of the small box is depressed; if the pressure decreases, the top of the box rises. The movement of the top of the box is transmitted to a needle on the dial of the instrument. Look at Fig. 13-5B and make sure that you understand how each part works. The common type of altimeter that is used in aircraft to indicate altitude is essentially nothing more than an aneroid barometer.

Fig. 13-5A. An aneroid barometer. The scale indicates the pressure in inches of mercury. The light-colored pointer indicates the pressure. The black pointer is set by hand over the light one by the knob in the center so the pressure change can be easily seen.



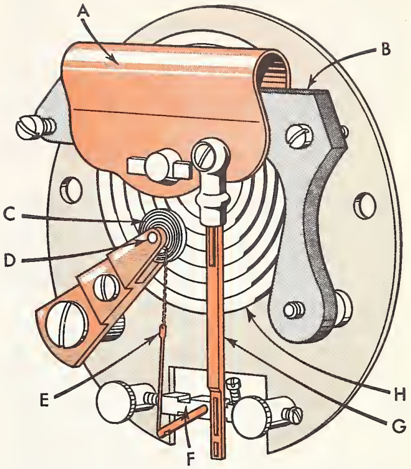


Fig. 13-5B. The mechanism of an aneroid barometer. The top of the hollow evacuated box, H, moves up and down with changes in pressure. The box is kept from collapsing by the piece of spring metal, A, fastened to the base, B. The arm, G, fastened to A moves about a pivot, and the arm F, turned by the pivot, moves the chain, E, back and forth. This chain is wrapped around the axle, D, and held taut by the hairspring, C. A pointer (not shown) fastened to the end of the axle, D, moves across a dial (not shown) to show the rise and fall of the pressure. The combination of arms, pivots, and levers is necessary to amplify the motion of the box and to give an easily visible movement of the pointer for small pressure changes.

Relative Humidity and Dew Point. In all weather stations the relative humidity and the dew point are determined by means of a table from readings of a wet- and dry-bulb thermometer. The dew point is important in forecasting fog. Why? Let us suppose that an aircraft takes off in the late afternoon on a clear day for an airport several hours' flight away, where the temperature is 80° F and the dew point is 76° F at the time he starts. He should know from these data that there is a strong possibility that fog will develop over the field before he arrives there. Why is this so? This is because the air will cool off as evening approaches. When it cools to 76° F the water vapor in the air will begin to condense and form fog, and the visibility may be reduced to the point where the flier will find difficulty in landing safely.

THE WEATHER MAP

Isobars and Wind. Fig. 13-6 is a surface weather map drawn from information received from the midnight Greenwich Mean Time (5 P.M. Mountain Standard Time) observations. This kind of map is published daily in many newspapers. It is a simplified map. It does not contain as much information as the official weather maps used by meteorologists, but it shows the most important features of the weather at ground level throughout the country.

Notice that this map shows many black lines drawn as irregular loops and circles, like contour lines. These lines are called ***isobars,*** and they indicate points of equal barometric pressure. Stated another way, ***all places on the same isobar have the same barometric pressure.*** The pressure is given in millibars. The millibar is the unit of pressure commonly used by meteorologists; ***1000 millibars are equal to 29.53 inches of mercury, which is approximately normal atmospheric pressure at sea level.*** For example, look at the isobar just west of Vancouver. It runs off the map just south of Vancouver, and the figure 1024 at the end of the line means that the barometric pressure at every point along this line is 1024 millibars. What is the pressure in inches of mercury? Isobars make it easy to locate the highs and lows across the country. Look at the map and see how the isobars form irregular concentric circles about the highs and lows. Could you tell which are highs and which are lows without the words "high" and "low" printed on the map? Isobars also indicate wind direction because, due to the rotation of the earth, ***winds blow nearly parallel to the isobars. About a low they are counterclockwise and slightly inward toward the center, and about a high they are clockwise and slightly outward away from the center.*** The isobars also give some indication of wind velocities. If the isobars are close together, wind velocities are usually high. If they are far apart, wind velocities are low.

You have probably already noticed that there is a small circle on the map at each place where observations of weather conditions are made, and that wind direction is indicated by the shaft from each station circle. The direction of the wind is along this shaft toward the circle. Barbs on the staff indicate wind velocity on the Beaufort Scale. A staff without a barb means a wind of 1 on the Beaufort Scale; a staff with half a barb, 2 on the scale; 1½ barbs, 4 on the scale; and so on. Look at the legend on page 128 which converts the numbers of the barbs into wind velocity. What Beaufort number and wind velocity are represented by three whole barbs? Can you locate a station on the map with this velocity? Can you find a place where a north wind is blowing?

Temperature, Clouds, and Precipitation. How are temperature, clouds and precipitation shown on a weather map? When the weather at a given station is clear, the station circle is unshaded. When it is partly cloudy, the circle is half shaded. How do you suppose a completely overcast sky is indicated? Rain, snow, and fog are represented by the symbols to the left of the station circles: snow; rain; drizzle; fog; thunder; lightning; thunderstorm; and a crosshatched area an area of continuous precipitation.

Can you read a weather map? Let us examine the Winnipeg station circle. Here we see a black circle, which means overcast, but no rain or snow. The temperature is 32°F, shown to the upper left of the station circle. The wind direction

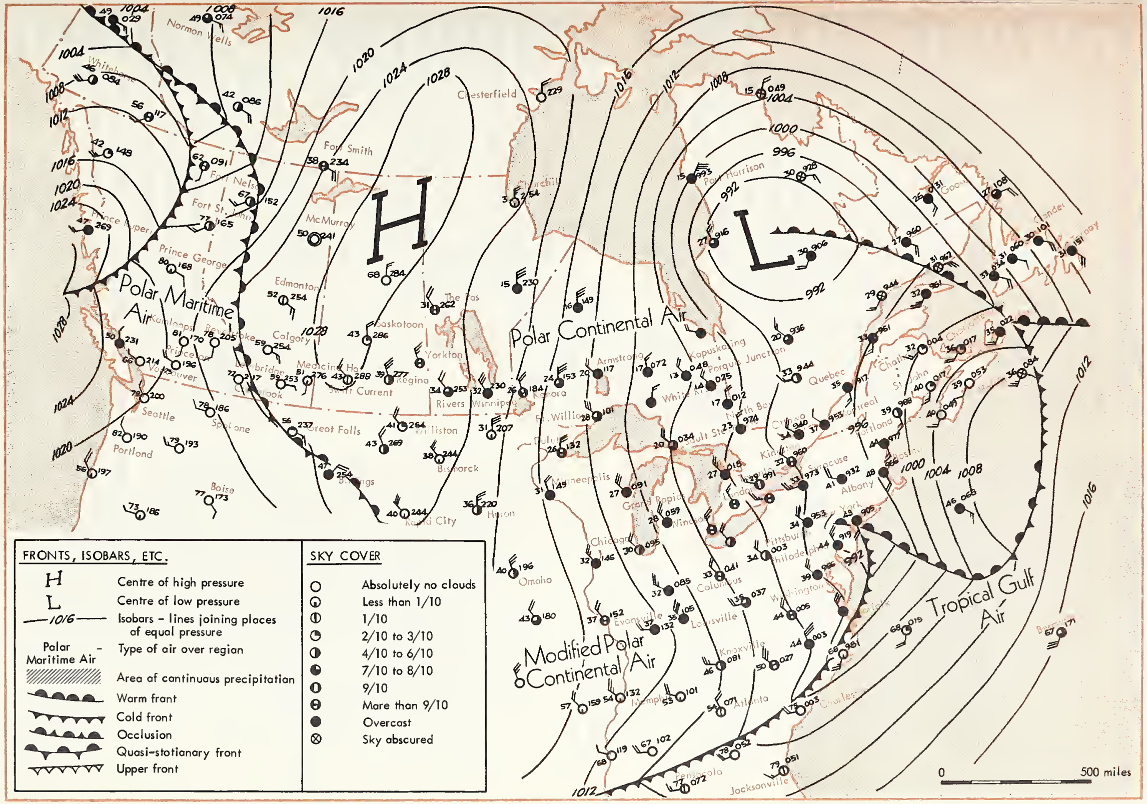


Fig. 13-6. A weather map of the type printed daily in the newspapers, showing weather conditions for a particular day. Temperatures are in degrees F. Each circle represents an observation station.

is northwest, and, since there are three whole barbs on the shaft, the wind velocity is between 32 and 37 miles per hour. The barometric pressure is shown to the upper right of the station circle as 230. This means the barometric pressure is 1023.0 millibars. When the pressure is below 1000 millibars, as at Quebec, three figures are given in the same way, so the reading of 957 on the weather map represents a pressure of 995.7 millibars.

Fronts and Air Masses. Fronts are shown on the map by long lines and markings along them. Along the cold front the markings are pointed. The markings are along that side of the front line toward which the front is moving. The front stretching from Philadelphia southward, into the Atlantic, is a cold front. You will notice, on the long front stretching from the Yukon Territory southward to Mexico, that the warm and cold front markings appear on opposite sides of the line. What kind of front is this, and what does it indicate?

The large air mass covering the prairie provinces and the central and eastern United States is a cold air mass that has been pushed down from the Arctic. One or two days before this map was made, it was northwest of the present location. It has become somewhat warmer on its way southeastward, but it is still cold enough to bring freezing temperatures as far south as the Great Lakes. In Churchill, on Hudson Bay, the temperature is only 3° above zero. The area of highest barometric pressure in this air mass is centered to the east of McMurray. It is called an anticyclone. You can see by the wind shafts on the station circles in this anticyclonic region that, with a few exceptions due to local conditions, the air is spiralling outward in a clockwise direction from the center of high pressure. The temperatures in the western half of this air mass are generally higher than those due east in the other half. The reason for this is that the air in the western half comes from the south, where is has been warmed, while the air in the eastern half comes from colder regions to the north.

Suppose we take an imaginary flight from Calgary, in the cold air mass, where the temperature is 59° and the wind is east-southeast, to Revelstoke, B.C. About half way there we would cross a stationary front and the clear dry air would change to slightly warmer hazier air. At Revelstoke, we would find the temperature higher and the air not quite so clear as in Calgary. We would also find that the wind would be blowing from the south. We would be in another air mass, a warmer one, which originated over the Pacific Ocean. If we flew to Fort St. John we would find, shortly after our arrival, that the wind would become gusty and would shift from southwest to west or northwest, and that the temperature would drop rapidly 20° to 30°. This would be the result of the eastward movement of the cold front that is shown approaching from the west.

The weather map also shows three cyclones. We can see only a small portion of one in the upper left-hand corner. It can be recognized by the low pressure in the Yukon. South of New York, however, we see a typical cyclone in its early stages, with a low pressure at its center. The cold front has just started to overtake the warm front, and a small occluded front is shown. Such a front (there are three in Fig. 13-6) is indicated on weather maps by alternating warm and cold front markings on the same side of the front line.

In examining a weather map you must always keep in mind that it shows the positions of highs, lows, storms and fronts at only one instant of time in the recent past. Remember that the weather is always changing; highs and lows move across the country, intensifying or weakening as they move; fronts appear and disappear and storms develop and die out. Looking at a weather map is like stopping a movie projector and "freezing" the motion on the screen by looking at only one picture. The sense of motion and change is lost. So in reading a weather map, a knowledge of how storms and fronts grow and die will enable you to estimate what changes have probably occurred between the time the readings for the map were taken and the time you examine it. It also enables you to visualize what changes in the weather are likely to take place in the near future.

Things to Remember

Weather data are collected simultaneously at many scattered stations throughout the country and telegraphed to the various forecast offices where weather maps are made from these data.

Weather maps give a complete picture of the weather over a large area.

Isobars are lines on the weather map connecting points that have equal barometric pressure at the same time.

Wind velocities are given on the weather map according to the ***Beaufort scale,*** 0 (calm) to 12 (hurricane).

The most important ***weather elements*** given on maps are precipitation, wind direction and velocity, barometric pressure, temperature, dew point and clouds.

Fronts, drawn on the map by the forecaster, show the movements of air masses.

Questions

GROUP A

1. What is (*a*) an anemometer, (*b*) an aneroid barometer?
2. How is the dew point measured?
3. Describe the construction and operation of a rain gauge.
4. Point out the following things on the weather map, Fig. 13-6: (*a*) cold front, (*b*) warm front, (*c*) isobar, (*d*) anticyclone, (*e*) station circle, (*f*) a place where the weather is clear, (*g*) a low, (*h*) a place where the weather is cloudy.

GROUP B

5. On the weather map in Fig. 13-6 locate the following: (a) cold air mass, (*b*) warm air mass, (c) two places where the atmospheric pressure is approximately 1025 millibars, (d) stationary front, (e) a place where the wind direction is south, (*f*) a place where the wind velocity is between 15 and 20 miles per hour, (g) a place where it is partly cloudy, (*h*) a cyclone.
6. From the weather map in Fig. 13-6 try to forecast what the weather will be 24 hours later in the following cities: (a) Regina, (*b*) Yarmouth, Nova Scotia.
7. How does an aneroid barometer work?
8. The mercury barometer reads 29.95 inches. Express this pressure in millibars.
9. How are weather data collected to make up the daily weather map?
10. How did Benjamin Franklin discover that storms are large whirlpools of air?

Things to Do

How to make a sensitive rain gauge · A simple sensitive rain gauge can be made from a tennis ball can and a cheap, 8-inch diameter funnel. Place the funnel in the can in an open area away from the drip of trees. A measuring stick should be made and marked off to measure the rainfall. To mark the stick correctly, remember that the depth of water collected in the can will be more than the true rainfall by a factor equal to the ratio of the cross-section area of the funnel opening to the cross section area of the can.

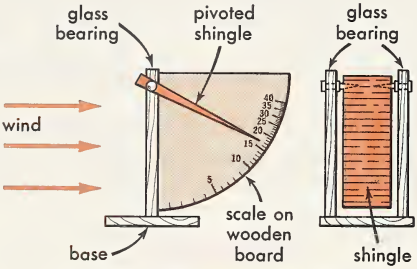


Fig. 13-7. A pressure plate anemometer.

How to make a pressure-plate anemometer · Mount a wooden shingle in bearings on a wooden frame as shown in Fig. 13-7. The bearings should be of short lengths of ¼-inch glass tubing. Place a small metal washer between the shingle and the bearing on each side. Use nails, driven into the thick end of the shingle for "axles." Calibrate the scale in miles per hour by holding the anemometer out the window of a moving automobile. The anemometer must always be held directly into the wind in a large open space as far above the ground as possible for accurate readings.

14

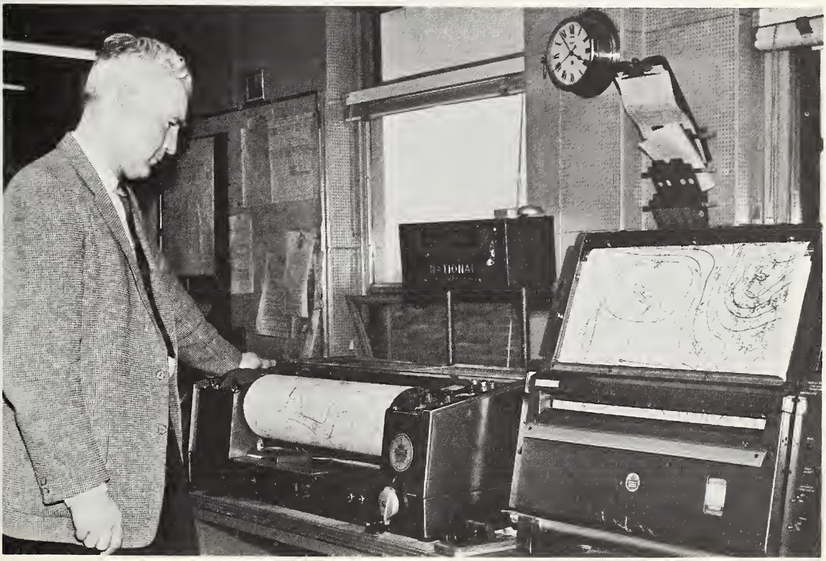
WEATHER FORECASTING (2)

THE UPPER ATMOSPHERE AND LONG-RANGE FORECASTING

The airlines and the Air Force pilots consult with the meteorologists to determine flying conditions along their flight paths before they start their flights. They also receive reports by radio on weather conditions along their flight routes.

Did you know that aircraft pilots often plan their flight routes to take advantage of favorable winds just as sea captains did in the days of sailing ships? On

Fig. 14-1. A Department of Transport meteorologist sending regional weather data, and receiving national data, in the form of weather maps transmitted by wire by means of facsimile machines.



transoceanic flights, planes do not always fly by the shortest geographical route; they can save both time and fuel by what is called ***pressure-pattern flying***. Before taking off, the pilot examines the pattern of the isobars on the weather map. By flying along the proper line of the isobars that surround anticyclones and cyclones along his route, he has a tail wind to aid him.

For example, across the North Atlantic, going from Montreal to London, he should fly around the ***north*** side of an anticyclone if possible. On the other hand, flying from London to Montreal he should keep to the north of the center of a low. Why is this? Draw a sketch of an anticyclone and its winds and imagine that you are going to fly a plane through it from west to east. Draw two possible flight paths—one north of the center and the other south of the center. Now can you see which path will give you the greatest benefit from tail winds?

Air Mass Analysis. Most modern weather forecasting is done by means of ***air mass analysis***. What does this mean? It means that a meteorologist tries to find out as much as he can, not only about the ***kind*** of air masses that cover the country at a given time but also about the "personality" of each air mass. For example if a tropical maritime (mT) air mass is approaching his area, he asks "Is it very damp, or only slightly damp? What is its temperature? Does it get cold rapidly above the ground or only slowly?" The answers to these questions will enable him to make accurate forecasts. But to get these answers he cannot depend upon observations made at ground level. He must also know the temperature, humidity and wind at every thousand feet or so through the air mass

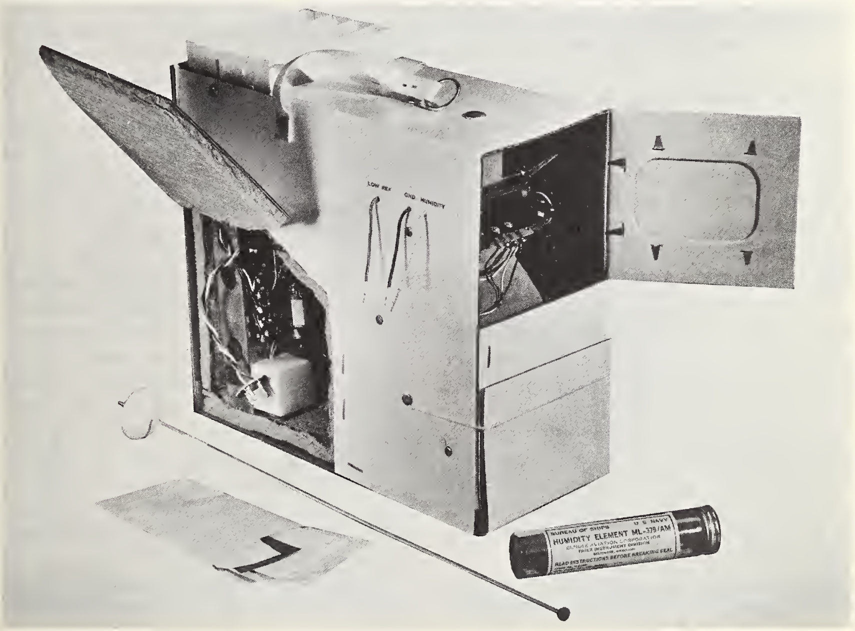
Fig. 14-2. Trans-Canada Air Lines pilots consulting with a meteorologist about the kind of weather they may expect along their flight route.

up to as high an altitude as possible. Why? Because at ground level all these factors are greatly changed by the nature of the earth's surface and they do not reveal the true personality of the air mass. For example, that part of a cold dry air mass that lies close to warm moist ground for some time tends to become warm and moist itself. In addition, the wind in this region will be reduced in velocity and changed somewhat in direction because of the friction of trees, hills, and buildings. At high altitudes, the air is little affected by the ground below.

How are such upper air measurements made? They are made by means of an ingenious device called a ***radiosonde*** (a French word meaning ***radio-soundings***) by which soundings of the atmosphere are made by radio. The radiosonde is a small box, about the size of a shoe box, containing a thermometer, barometer, and relative humidity indicator. Also included in the box is a tiny radio transmitter and batteries to power it. The thermometer, barometer and humidity indicator are ingeniously hooked up to the radio transmitter so that, as the radiosonde rises, it transmits the temperature, pressure, and humidity back to the earth's surface by means of coded signals.

The radiosonde is lifted through the atmosphere to altitudes above 70,000 ft. (about 12 miles) by a balloon. During its ascent it is followed by radar which indicates the direction and velocity of its drift at different altitudes. Since the balloon drifts with the wind, the meteorologist knows the wind velocity and direction at different altitudes. The balloon floats upwards for about 90 minutes until it reaches its "ceiling." Here it may

Fig. 14-3. A radiosonde. Part of the barograph for measuring pressure can be seen in the upper right. The lower left compartment houses the tiny radio transmitter. The antenna and a capsule containing the humidity measuring device are in the foreground.


LOW REF
GND
HUMIDITY
BUREAU OF SHIPS
U S NAVY
HUMIDITY ELEMENT ML-379/AM
READ INSTRUCTIONS BEFORE BREAKING SEAL

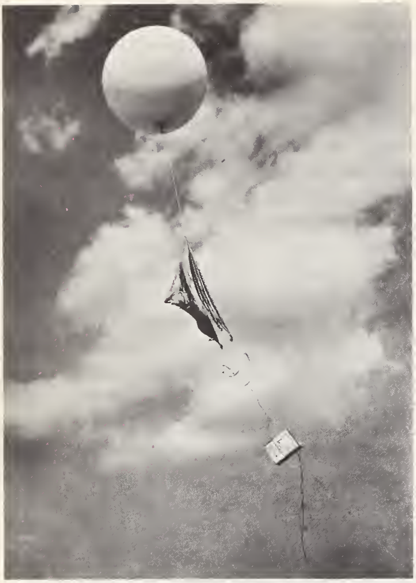


Fig. 14-4. A radiosonde rising through the atmosphere. Between the helium filled balloon and the radiosonde is a deflated parachute that opens when the balloon breaks at high altitude. In this type of radiosonde the antenna is a short wire hanging beneath the radiosonde.

burst or it may begin to fall slowly as the gas diffuses out of the balloon. Attached to the radiosonde is a small parachute which gently lowers the radiosonde to earth. It may be that you will find one of these instruments that has served its usefulness to the weather service. These may be retained by the finder.

Radiosonde data are distributed by teletype and used to make weather maps of the air at different levels above the ground. These maps are used to supplement the ground-level maps.

Long-Range Forecasting. By using a series of weather maps like the one given in Fig. 13-6, a weather forecaster is able to estimate, with fair accuracy, the paths of air masses and storm centers for the coming 24 to 48 hours. With these estimates, he then is in a position to forecast the weather for any particular locality for the coming two days. When he tries to use the same method for forecasting beyond the third day, he finds considerable difficulty. A storm center dies, or at least becomes much less active. Another storm center is born in the manner shown in Fig. 12-9. A flow of mild air changes its direction. Because of these and other difficulties, he finds that daily forecasts become unreliable for periods beyond two or three days ahead.

To make forecasts for periods beyond three days, the forecasters make use of mean weather maps. They have learned that certain features of the weather maps which may not show very clearly on the daily charts can be found on these mean charts. Fig. 14-5 shows a mean wind flow for a month at 10,000 feet above mean sea level. The map shows that on the average for the month there are southwest winds over the Pacific, a "ridge line" with west winds over the western section of the continent. Over the Great Plains eastward to Hudson Bay and the Mississippi the air is moving somewhat southward. A "trough line" is found lying on a line from Hudson Strait to Lake Huron, and east of the trough line the winds are southwest. At the center of such a stream of air, at about 35,000 feet, there is a jet of very high-speed air, called the jet stream, where the wind may reach a velocity as high as 300 miles per hour. Locating the path of the jet stream therefore helps to locate the movements of the upper air.

As seen in Figs. 14-6A and 6B, along the mean ridge line the temperatures tend to be warmer than normal and the precipitation is light. Low pressure centers tend to become more intense in mean trough lines and move northeastward, giving heavy precipitation. The area of heavy precipitation can be observed in Fig. 14-6A to extend from the Gulf of Mexico coast northeastward to Quebec and Labrador. Can you relate other parts of Figs. 14-6A and 6B to the mean wind flow shown on these maps?

These mean troughs and ridges move slowly, generally but not always eastward.

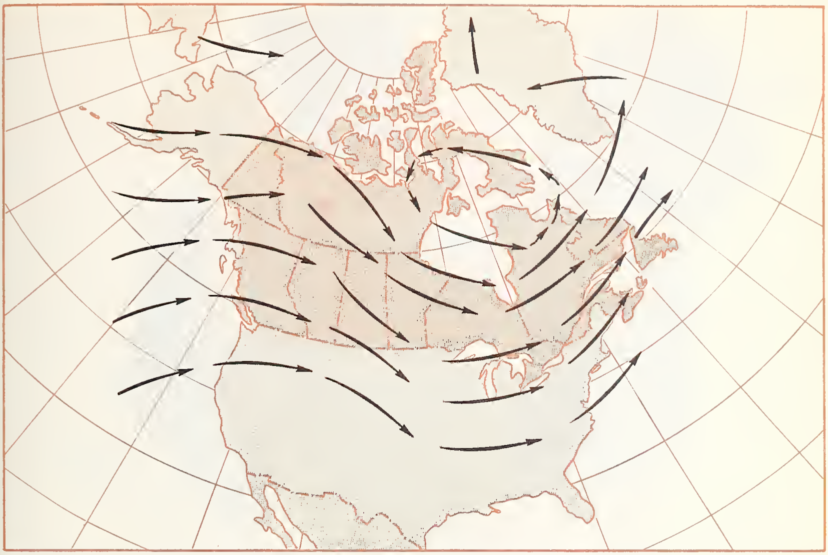
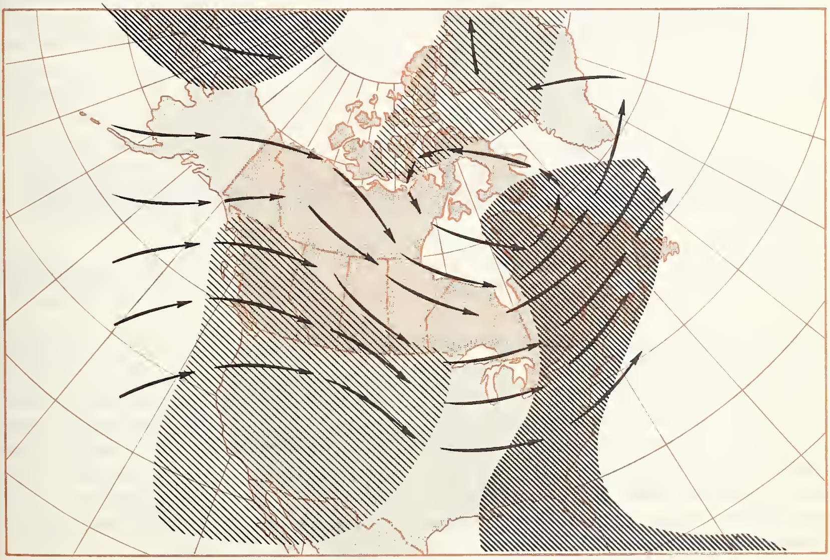


Fig. 14-5. The air flow in the atmosphere at an altitude of 10,000 feet.

Fig. 14-6A. Precipitation resulting from the high altitude air flow shown in Fig. 14-5. Heavy shading shows heavy precipitation, no shading shows moderate precipitation and light shading shows slight precipitation.



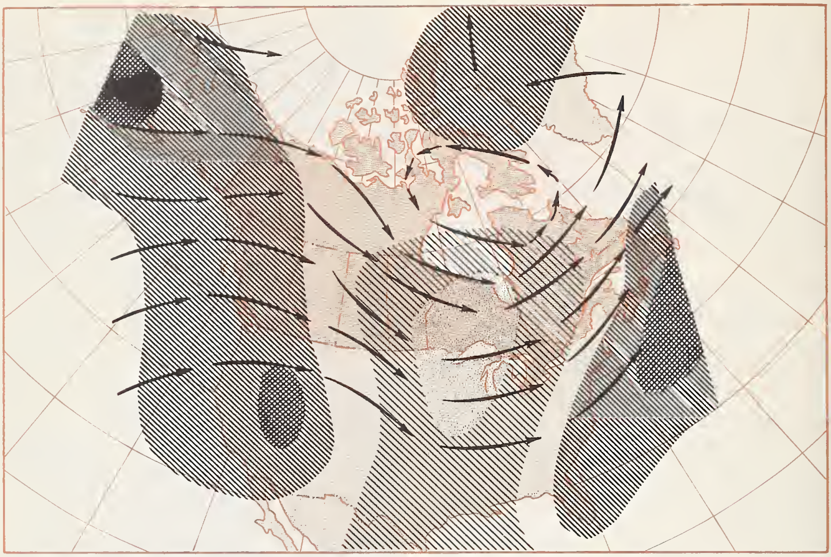


Fig. 14-6B. Surface temperatures resulting from the high-altitude air flow shown in Fig. 14-5. The shaded areas indicate temperatures as follows: very heavy, much above normal; heavy, above normal; no shading, normal; light, below normal; very light, much below normal.

By studying these movements carefully, the U.S. Weather Bureau, Extended Forecast Division, has been able to make forecasts for the succeeding month. These do not attempt to tell the weather for the individual days, but give a generalized forecast for the month as a whole. Can you think of any group of people who would find such a generalized forecast useful? These forecasts are sometimes distributed in the form of charts like those seen in Figs. 14-6A and 6B.

Occasionally one hears or reads of weather forecasts of daily weather for periods a week or more in advance. These forecasts are based upon the average weather for the season. Thus they would forecast showery weather with thunderstorms in the spring for most of Alberta, but with less shower activity in southern Alberta as the season advances. By emphasizing the average weather, they tell those who read them what they should expect. But because they are based upon averages they are equally good for the same season of any year.

The Troposphere and the Stratosphere. The lowest layer of the atmosphere is called the ***troposphere.*** It is a relatively thin layer, about 7 miles (or 40,000 feet) in depth, into which is packed nearly 80 per cent of the gases of the atmosphere. It is in this extremely thin shell that all our weather takes place. The "jet stream" is a river of wind in the troposphere. It encircles the earth, blowing from west to east in the Northern Hemisphere all the year round.

The next layer is the ***stratosphere,*** which extends above the troposphere for a distance of 50 miles or so. The stratosphere, which contains only 20 per cent of the atmosphere, is a region where a temperature inversion occurs. At the lower levels of the stratosphere the temperature is about -69°F, but it then increases with increased altitude until at about 30 miles the temperature is well above freezing. It drops again to about -100°F at about 50 miles. In the stratosphere there are no

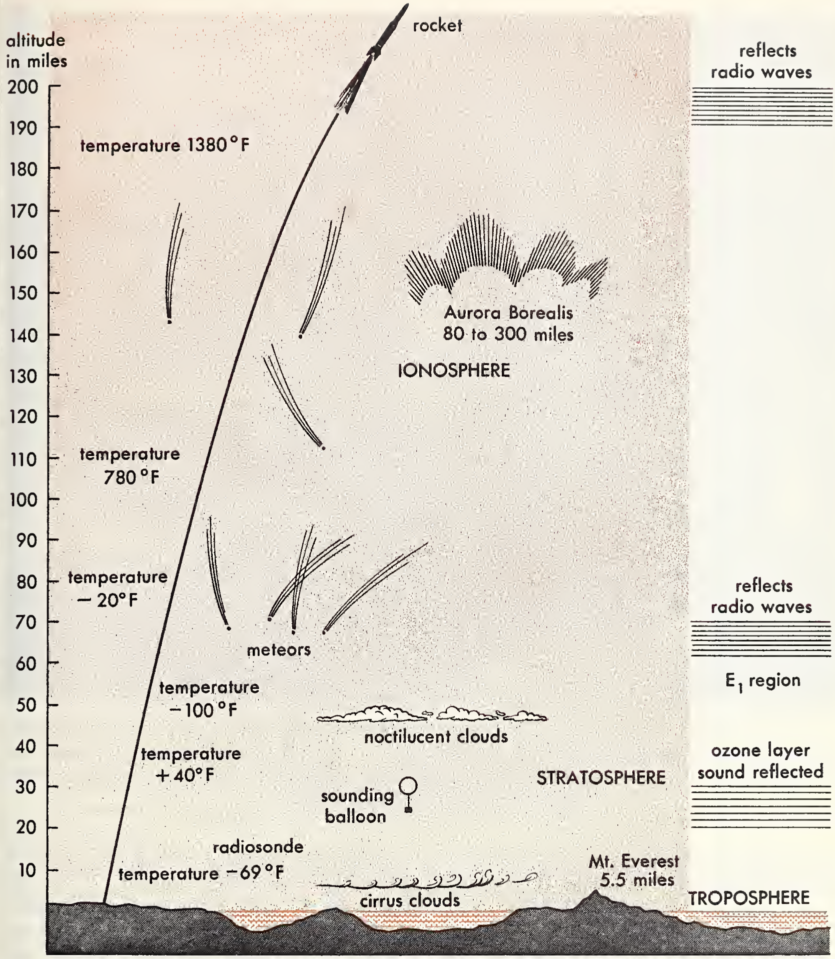


Fig. 14-7. A vertical cross section of the atmosphere. Storms, fronts, and rain clouds all occur in the troposphere. Above the troposphere is the stratosphere, a layer of low temperature. In the ionosphere, which reflects radio waves, there are very few air molecules and the temperature becomes very high.

convection currents, no well-defined fronts; in short, no weather.

Men have penetrated into the lower parts of the stratosphere in planes and balloons. This is an ideal altitude for aircraft, since it is above all bad weather. Moreover, aircraft flying near the bottom of the stratosphere can, in certain cases, take advantage of very high-speed tail winds by flying in the jet stream. More and more planes are being designed to fly in the stratosphere, but such planes must of course have pressurized cabins. Why?

Earth Satellites. Above the stratosphere is a third layer called the ***ionosphere,*** which reflects radio waves, extending outward for several hundred miles. Scientists

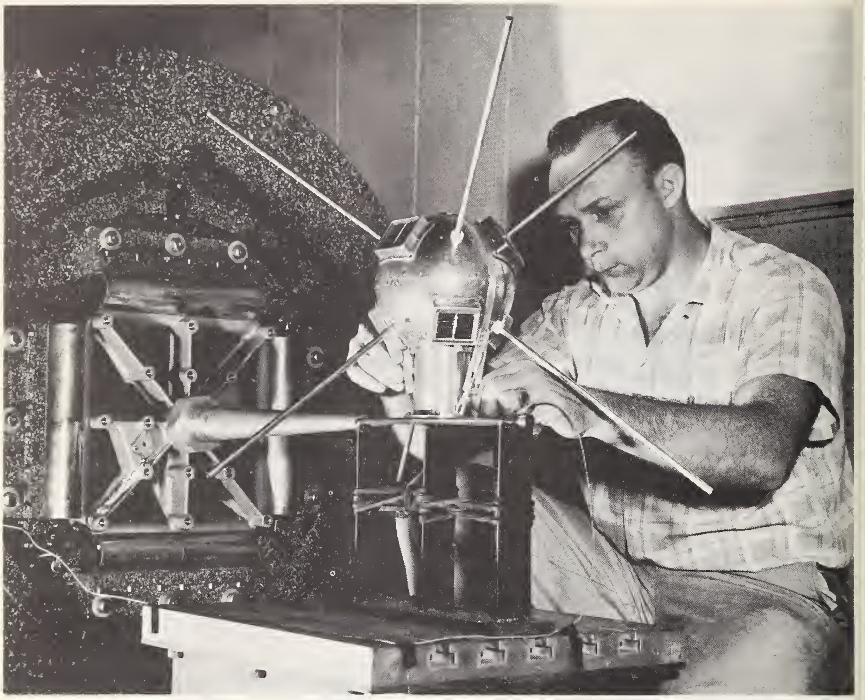


Fig. 14-8. A small artificial satellite. Notice the six radio antennas and also the windows to admit sunlight required to power solar batteries. Such satellites can circle the earth in less than two hours at altitudes above 300 miles, transmitting information about temperature, meteor impacts, and cosmic rays to receivers at the earth's surface.

have learned something about the atmosphere a hundred miles high by means of high-altitude rockets which are sent up carrying radiosonde instruments. But such rockets are in the air for only a few minutes before falling back to earth. Much more information can now easily be obtained from earth satellites. These satellites contain instruments and a radio transmitter like a radiosonde and travel at a very high speed in an orbit about the earth, like the moon only much closer to the earth. The satellites are small, only a few feet in diameter and, eventually, due to friction, they spiral back to earth. In doing so they burn up like meteors as they reach the lower and denser atmospheric layers.

The instruments in a satellite can transmit to the ground a wealth of valuable information about cosmic ray intensity, the number and sizes of the tiny meteors that drift in space, and the intensity and nature of the sun's radiation. None of these can be measured accurately at ground level because of the blanketing effect of our atmosphere. Satellite instruments can also tell us about the temperatures in the ionosphere, the earth's magnetic and electric fields at these high altitudes and how radio signals (those from the satellite itself) penetrate the ionosphere. Slight changes in the orbit of the satellite give information about the shape and size of the earth's equatorial bulge, while the change in speed and the life of the satellite tell us about the density of the air several hundred miles up. More will probably be learned about the causes of the jet stream and its changing path, and better weather forecasting will result.

The placing of the first man-made satellite in its orbit in 1957 was one of the great landmarks in the history of science. It represents a culmination of much of man's search for knowledge, for beneath the technology and mechanics of the rockets that lift the satellite to its orbit lie all of the fundamental laws of physical science—laws whose understanding began with Copernicus, Galileo, Newton, Lavoisier and others. With the advent of the satellite, man is about to enter a new age—the age of the exploration of space.

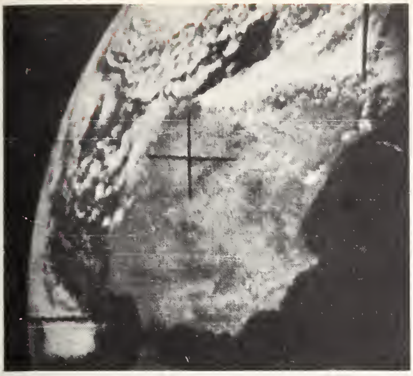


Fig. 14-9. Spain and the Straits of Gibraltar. Clouds cover the northern slopes of the Pyrenees, and the circular cloud pattern of a cyclonic storm shows prominently in the north Atlantic. This picture was taken from Tiros III Weather Satellite on July 15, 1961.

Things to Remember

Flying around cyclones and anticyclones so as to take advantage of tail winds is called ***pressure-pattern flying.***

Modern weather forecasting is greatly aided by ***air mass analysis*** which requires a knowledge of temperature, humidity, pressure and wind at high altitudes.

A ***radiosonde*** carried aloft by a balloon transmits automatically by radio the temperature, pressure and humidity at different altitudes.

Long-range forecasting is accomplished by observing the direction of airflow at high altitudes.

Long-range forecasting can predict changes in ***average*** temperature, precipitation and number of storms for periods of three to four weeks ahead.

In the ***troposphere,*** temperature usually decreases with altitude.

The ***stratosphere*** begins at about 40,000 feet.

Questions

GROUP A

1. What is pressure-pattern flying?
2. What is air mass analysis?
3. What is a radiosonde and what is it used for?
4. What is long-range forecasting? Of what value is it?
5. What is the jet stream?
6. What is (*a*) the troposphere, (*b*) the stratosphere?

GROUP B

7. Long-range forecasting cannot predict whether there will be a storm or clear weather on any given date three weeks in advance. Why is this? In view of this, is long-range forecasting of any value? Explain.
8. Some almanacs predict the exact weather for a given day a year in advance. Do you think they are of any value? Explain.
9. What does the jet stream have to do with long-range forecasting?
10. What is the difference between an earth satellite and a radiosonde? How are they similar?
11. Why should a trans-Atlantic airplane fly ***south*** of a cyclone in order to make the best time when flying from Montreal to London?

Things to Do

To organize a meteorological club · Start a meteorological club in your school. You can obtain simple weather instruments from any reliable instrument company or you may make them yourself. Make daily or twice daily observations at a fixed time of the following weather elements: (a) air temperature, (*b*) maximum daily temperature, (c) minimum daily temperature, (*d*) kinds and amounts of clouds, (e) state of weather (clear, partly cloudy, rain, fog, etc.), (*f*) barometric pressure, (g) change in pressure (rising or falling), (*h*) wind direction and velocity, (*i*) amount of precipitation since the last observation.

The members of the club can take turns taking the readings, and weekly meetings can be held to discuss the weather during the past week and to compare the club's observations with those published. The daily weather map is published in most newspapers. Clouds can be observed and identified at the weekly meetings, and various aspects of meteorology discussed. The daily weather observations can be kept in a permanent file for reference at future meetings.

15

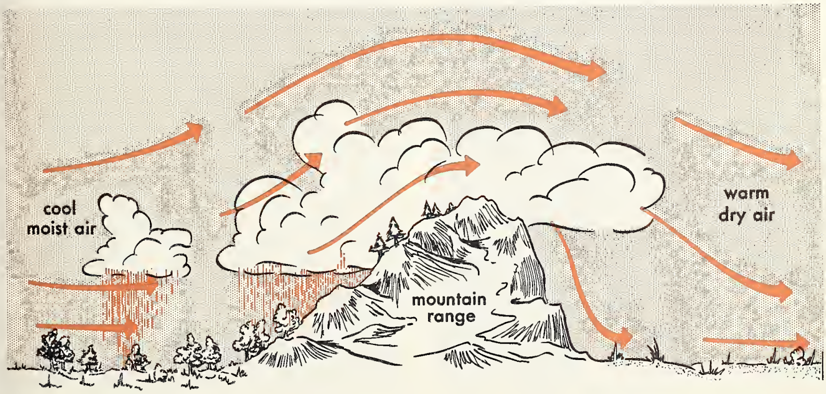
WHAT IS CLIMATE?

When we speak of the ***weather,*** we mean the day-by-day changes of temperature, rainfall, wind, clouds, humidity, and so on. ***Climate,*** on the other hand, refers to the average of these weather conditions in an area over a period of years. For example, if a warm front over Saskatchewan produces rain for several days, we say the ***weather*** is wet. But because rain occurs infrequently in Saskatchewan, and the total amount of rainfall each year is small, we say the ***climate*** of Saskatchewan is dry.

As one might expect, the warmest climates are found near the Equator and the coldest near the poles. Small islands, being entirely surrounded by water, have damp climates, whereas areas within a large continent, far from large bodies of water, generally have dry climates. Western China and Siberia, for example, have dry climates. The climate of any region is determined by such factors as the direction of its prevailing wind, its nearness to the Equator or poles, its nearness to large bodies of water, and the location of high mountain ranges.

Mountains as a Cause of Rain. Rain and snow fall only when a mass of moist air rises, and expands and cools enough to cause condensation. This can occur not only along fronts, but also when moving air is forced up over a mountain range.

Fig. 15-1. Rain falls when warm moist air, forced to rise over a mountain range, expands and cools. However, the air gains heat from the condensation of the water vapor, and as it descends on the far side of the mountains, it is warm and dry.



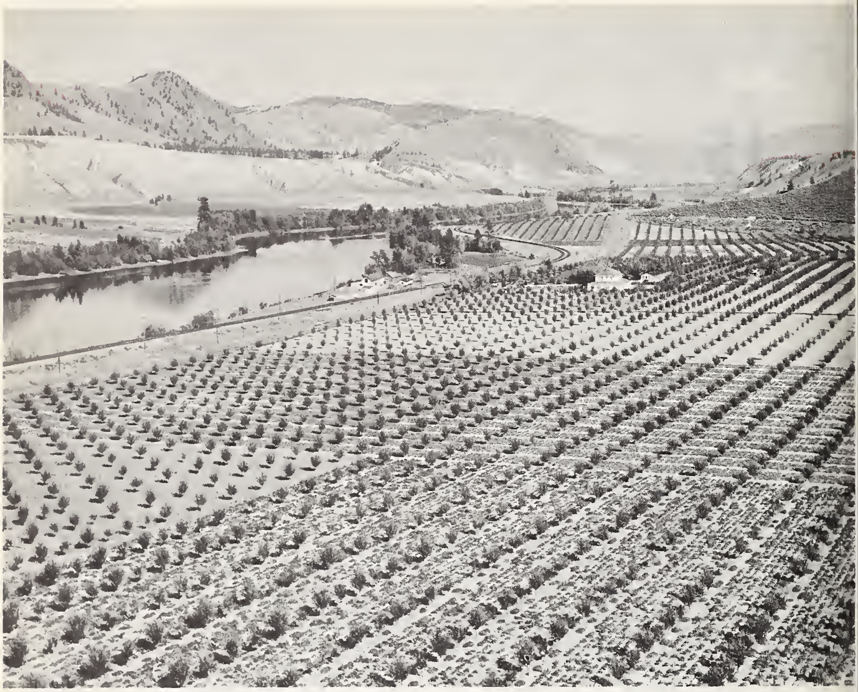


Fig. 15-2A. Apple orchards in the fertile South Thompson Valley, British Columbia.

Let us compare the climate of the western slopes of the Rocky Mountains with that of their eastern slopes. The prevailing winds that blow across British Columbia are westerly, bringing vast amounts of water vapor from the Pacific Ocean. This moisture-laden air is forced to rise up the mountain slopes as it flows over the various ranges. In rising it cools, so that much of its water vapor condenses, resulting in a large amount of rainfall on the western slopes. This rain-water runs into the valleys of British Columbia, providing moisture to irrigate the crops which then grow well in their fertile soil. By the time the air has passed to the eastern slopes it has lost much of its water vapor, and the latent heat of condensation has warmed it. Southern Alberta and southwest Saskatchewan as a result have sunny skies and light rainfall. An extreme example of this phenomenon is found at Death Valley, just east of the Sierra Nevadas, which is the driest and hottest place in the United States.

Monsoons. The ***monsoon*** is the name given to the seasonal prevailing winds in India. During the winter, the prevailing wind over the Himalayas in northern India is from the north. Since the area from which it blows is a large mass of land, this wind does not contain much moisture, and it is cool. In the summer, the prevailing wind is from the opposite direction; it comes from the south and carries large quantities of moisture from the Indian Ocean; it is a warm wind and makes India hot and humid during the summer.

The most outstanding feature of this summer monsoon is the large amount of rain that falls on the southern slopes of

Fig. 15-2B. Death Valley in southeastern California. Can you explain the difference between the climates by referring to Fig. 15-1?

the Himalayas, the highest mountain range in the world. The strong moisture-laden wind from the Indian Ocean blows toward the mountains, and the result is that nearly all the moisture is deposited among the foothills of the southern slopes. It rains continuously for a whole season, and more than 100 inches are usually recorded during each of the months of June, July and August. The heaviest rainfall in the world—nearly 500 inches a year—has been recorded there.

As one would expect, the climate north of the Himalayas is hot and dry in summer and cold and dry in winter. In this region is the famous Gobi Desert.

Maritime Climates and Continental Climates. The climates of the earth are roughly divided into ***maritime*** and ***continental.*** What is the difference between them? Land masses with prevailing winds from the ocean are said to have ***maritime climates.*** Such climates can be described as having small changes in temperature between summer and winter (or from day to day) and a rather high precipitation and relative humidity. This small range in temperature is a result of the high heat capacity of sea water, which neither heats up nor cools off rapidly.

In the temperate zone of the Northern Hemisphere, the land on the west coasts have a maritime climate. England also has a maritime climate, and so do small islands scattered over the oceans. On the other hand, the interior of large continents, and the east coasts in the temperate zone of the Northern Hemisphere experience continental climates. Since the prevailing wind is west, such areas receive few air masses that originate over water and so they are dry. Since dry earth has a low heat capacity they have large temperature differences between summer and winter. The low humidity of a continental climate often gives rise to deserts; the Sahara Desert in North Africa, the desert in Central Australia and the Gobi Desert of Siberia and China are all surrounded by large land areas.

The large temperature changes that occur in continental climates are well-illustrated by Verkhoyansk in Siberia where, from the table following we see that the average temperature for the coldest month is −58° F whereas for the warmest month the average is 56°F, a difference of 114° F between the coldest and warmest months. Indeed, one of the lowest temperatures ever recorded was at Verkhoyansk (−94° F). It is sometimes referred to as the "cold pole" of the world since it experiences some of the lowest temperatures of any spot on the globe. The lowest temperature ever recorded was slightly below −100° F at the South Pole. The differences between maritime and continental climates are shown in the table. It must be remembered, however, that summer and winter temperature differences are greater in high latitudes than near the Equator where the sun heats the earth almost equally the year round.

The "Stormy Westerlies" and the "Trades." The climates all around the world in the temperate zone, whether maritime or continental, are noted for the frequency of large storms and the variability of the weather. That is why they are called zones of "stormy westerly" winds. It is in these latitudes that warm and cold air masses interact along fronts to produce cyclonic storms.

On each side of the Equator where there are few fronts, the winds are steady trade winds, and the climate shows little variation in weather. The storms experienced here are mostly local thunderstorms due to convection, except for areas swept by seasonal hurricanes and rain caused by the flow of air over high mountain ranges.

The Climate of Canada. Canada has both maritime and continental climates. A good way of examining the climate in different parts of the country is to look closely at charts of wind direction, temperature and rainfall (Figs. 15-4, 15-5 and 15-6.)

Prevailing Wind. The prevailing winds at different places in Canada in winter and summer are shown in Fig. 15-4. You will notice that most of the arrows point eastward, indicating that the prevailing wind is the expected wind from the west. However, a portion of the east coast shows southwesterly winds in summer and northwesterly winds in winter. This seasonal change is, of course, a monsoon effect.

Average Temperature. The average

	Average Temp. of coldest month	Average Temp. of warmest month	Annual rainfall in inches	Type of climate
Alice Springs, Australia	53° F	83° F	10	Cont
Verkhoyansk, Siberia	−58	56	5	Cont
Dutch Harbor, Alaska	32	51	63	Mar
Edmonton, Canada	8	63	18	Cont
Midway Island (Pacific)	65	78	43	Mar
Denver, Colorado	29	73	14	Cont

Fig. 15-3. Climate at selected places around the world.

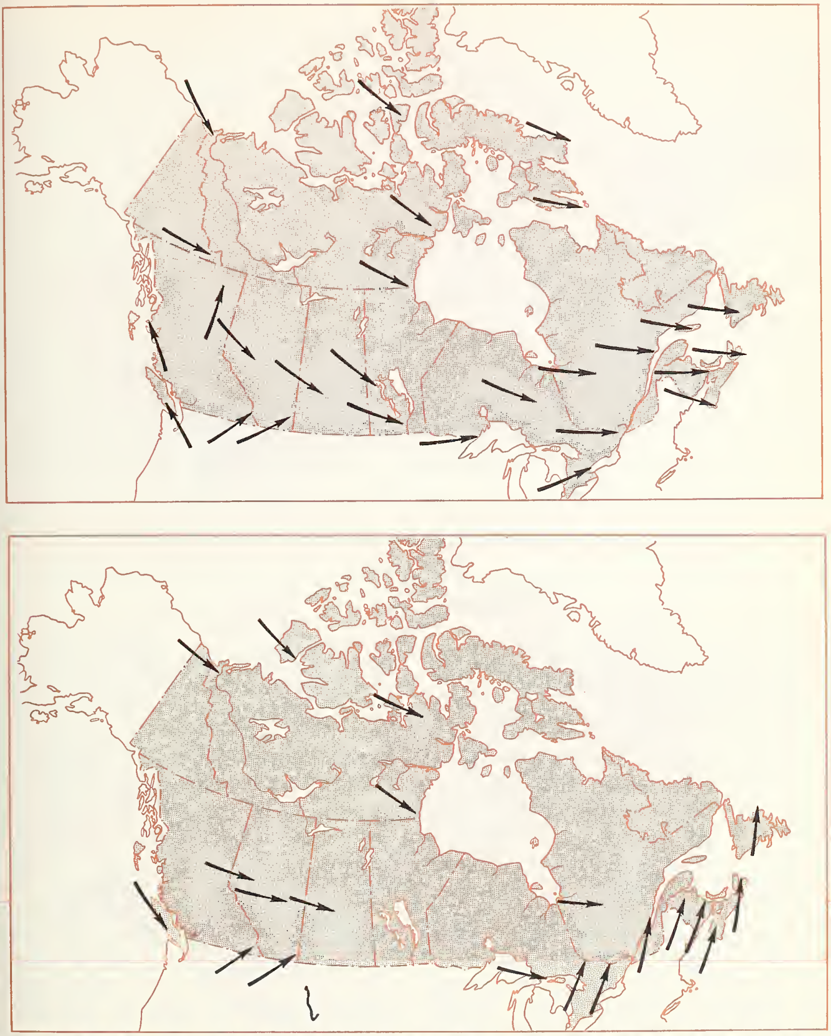


Fig. 15-4. The most probable wind directions in Canada in winter (top) and in summer (bottom).

daily temperatures throughout the country are shown in Fig. 15-5. It indicates that the east coast (continental climate) has a greater variation in temperature than the west coast (maritime climate) while the center of the country (continental climate) shows the greatest seasonal temperature variation of all.

Precipitation. In Fig. 15-6 we see that the maritime climate of the west coast has the greatest precipitation in wintertime, while the lowest precipitation occurs in the central part of the country (continental climate). How can you explain these facts?

Charts of average temperature, pre-

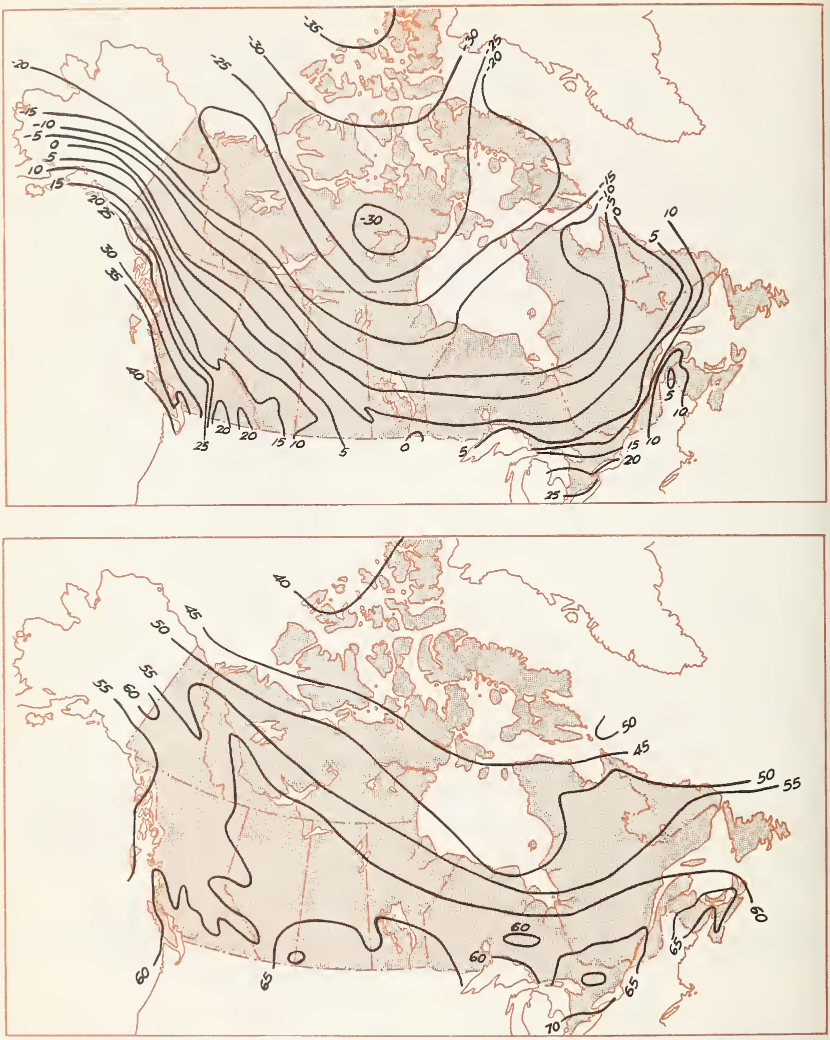


Fig. 15-5. January mean daily temperatures (top) and July mean daily temperatures (bottom) in Canada.

cipitation and wind direction, as well as charts of amounts of sunshine are very valuable to farmers and businessmen since they tell what can be expected in the way of climate at different places. For example, before building a large dam for flood control, irrigation and generating electric power, the amount of rainfall to be expected at the location of the dam site must be known. The size of the dam and reservoir, the amount of land that can be irrigated, and the amount of electric power that can be generated all depend on an accurate knowledge of the rainfall in the particular area.

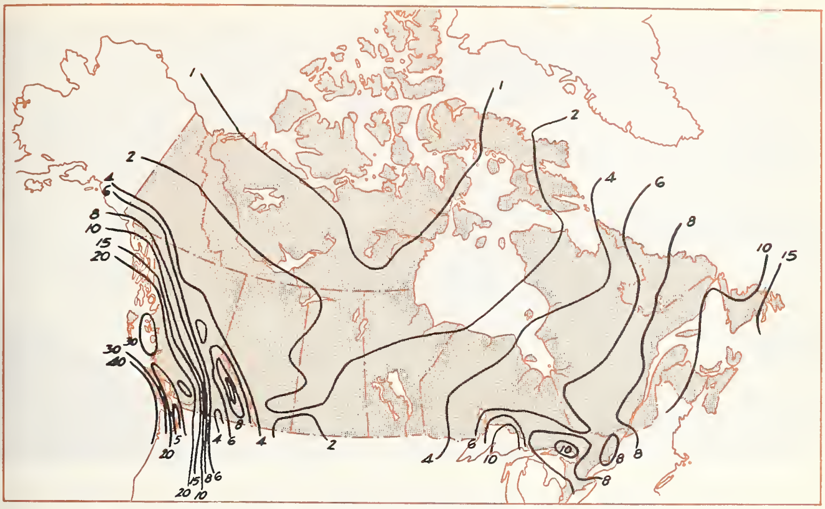


Fig. 15-6 Mean winter (December to February) precipitation in inches in Canada.

CONTROLLING THE WEATHER

Meteorologists and physicists sometimes dream of being able to control the weather, so that every part of the world may always have good weather. If the arctic and antarctic regions could be warmed so that food could be grown there, or if the heat of the tropical areas could be tempered to make them more livable, many of the social problems resulting from an increasing world population and an insufficient food supply would be solved. Today such control of the weather sounds rather fantastic, but progress is being made.

The first attempts to control weather were made by the ancient Egyptians. They practiced irrigation in an effort to moisten areas in which there was no rainfall. These early efforts to overcome natural deficiencies have today developed into irrigation projects such as those in southern Alberta and Saskatchewan.

Weather Control in Orchards. Another form of weather control is the use of smudge pots to prevent the freezing of fruit trees and blossoms in the orchards of British Columbia during a cold wave. At night when the weather is abnormally cool and the sky clear, smudge pots are put out in the orchards. Oil burned in these pots produces large amounts of smoke, forming a cloud over the orchard which prevents the escape of heat from the ground.

In some orchards freezing temperatures close to the ground are prevented by means of large electric fans placed on ladders between the trees. These fans circulate the air so that cold air near the ground is replaced by warmer air from above. These methods, coupled with timely warnings from the Vancouver Forecast Office, save many acres of fruit every year.

Fog Dispersal. Ever since the early days of flying, scientists have worked on methods of dispersing fog. Several methods have been used with very limited success. In one of these, large sound-generating machines are set up along the edge of a runway. When a plane is about to

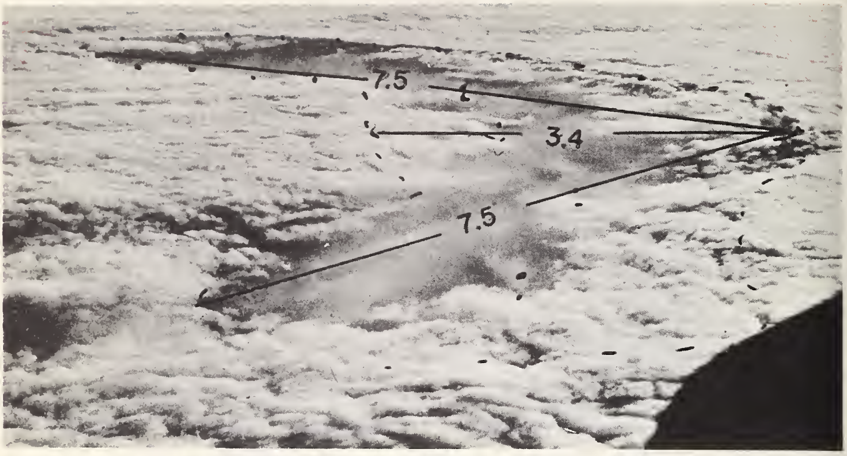


Fig. 15-7. This 15-mile long, 3-mile wide L-shaped hole, roughly twice the area of Manhattan Island, New York City, was produced in a supercooled or icing cloud by seeding the cloud with dry-ice pellets. Since such a hole is free of dangerous icing conditions, an airplane could descend or ascend through it safely.

land, the sound generators are turned on. They produce tremendous amounts of sound energy of a pitch too high to be audible to the human ear. The sound vibrations produced in the air knock the fog droplets together until they form drops large enough to fall to the ground as a light rain, leaving the runway clear of fog until the plane has safely landed.

Rain Making. Rain and snow have been produced by the use of dry ice. In one method, an airplane flies through a dense cloud scattering small pellets of dry ice. As a result, droplets in the cloud join to form large drops of water which fall to earth. A few pounds of dry ice may produce an increase in precipitation. In another method silver iodide is carried up from the ground in the convection currents from heated blowers. Obviously rain cannot be "made" unless there is moisture in the air. The dry ice or silver iodide merely acts as a "trigger" to start precipitation from a moisture-laden cloud that might not produce snow or rain if left alone.

Things to Remember

Climate is the average weather over a period of years.

Small land masses surrounded by water have a ***moist climate*** (maritime climate).

Large land masses, far from oceans have a ***dry climate*** (continental climate).

Continental climates have large temperature ranges.

Maritime climates have small temperature ranges.

A ***monsoon*** is a seasonal prevailing wind that blows from one direction part of the year and from the opposite direction the rest of the year.

Smudge pots are used to prevent freezing in orchards.

Precipitation can be started by scattering such substances as dry ice or silver iodide in moisture-laden clouds.

Questions

GROUP A

1. What is the difference between weather and climate?
2. (*a*) What is a monsoon? (*b*) Where does the best known monsoon occur?
3. Where does the world's greatest rainfall occur?
4. Where do the world's lowest temperatures occur?
5. What are smudge pots used for?
6. What is the difference between a continental climate and a maritime climate?
7. Describe what you think would be the climate of (*a*) Bermuda, (*b*) Ceylon.

GROUP B

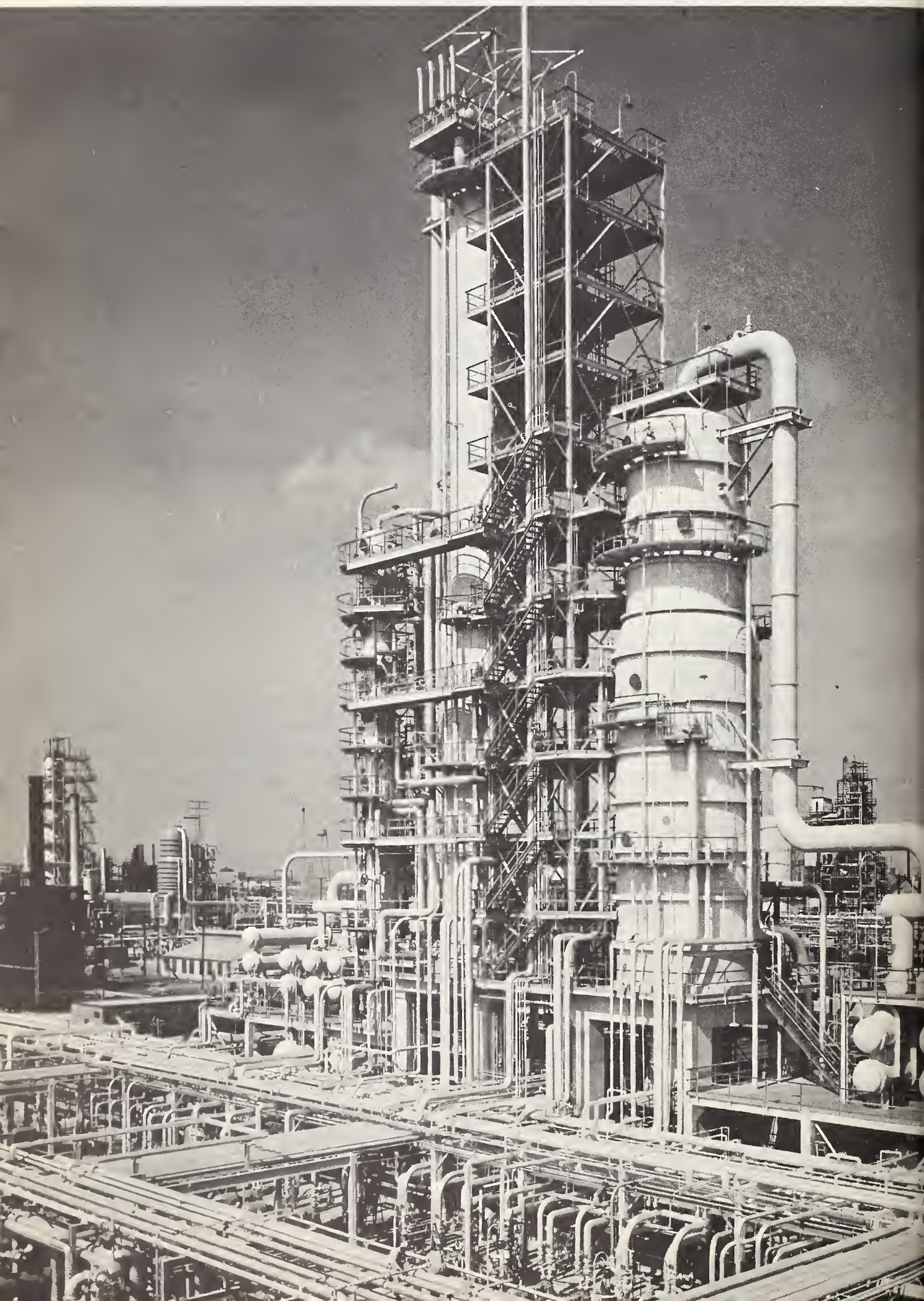
8. How do mountains cause rain?
9. Describe one method of dispersing fog.
10. How may precipitation be started in a cloud that is heavily saturated with water vapor?
11. Why is the area east of the Rocky Mountains dry?
12. Explain the causes of the heavy rains in India. Why do they not occur in winter?
13. What has caused the Gobi Desert?
14. Why does it never become very hot or very cold in Bermuda?
15. Why is it very hot in summer and very cold in winter in Winnipeg.

Things to Do

Determining microclimates · A microclimate is the average weather over a very small area—a few square miles or less. There can be quite noticeable differences in average temperature, minimum temperature, rainfall and wind direction between two places only a few hundred yards apart. This is one of the reasons why one field grows better crops than a neighboring field.

You can determine the microclimates of different parts of your town by measuring temperature, precipitation and wind direction at scattered points throughout the town. Each member of the class can make measurements at his own location and then the results can be compared. Be sure that everyone's instruments all read alike before they are set up where they are to be used.

Chemistry



LOOKING AHEAD

It has been said that civilization began when primitive man first learned to build a fire. With fire he was able to cook his food, fashion weapons and tools, protect himself against wild animals, and, in general, make life more comfortable.

But it was a far cry from making fire and knowing what happened when a fire was made. The cause of fire was a complete mystery to early man. It is not surprising, therefore, that he was superstitious and apprehensive. When a flash of lightning started a forest fire he fled in terror, believing it to be the evil work of an angry fire god intent upon destroying him.

Fire remained a mystery through the centuries; its secret was not revealed until the end of the 18th century. It was Lavoisier, a famous French chemist, who found the explanation. Fire, he said, is caused by a chemical reaction between a combustible substance and the oxygen of the air.

An understanding of fire led to new ways of transportation. First the steam locomotive used coal as a fuel and later the internal combustion engine used gasoline. With the advent of the airplane the demands for gasoline were insatiable. A cracking process was discovered which doubled the gasoline obtained by distilling crude oil. The byproducts of cracking opened up new fields that became a chemist's paradise. Out of small molecules he synthesized a vast array of larger ones and these he fashioned into plastics, textiles, drugs, perfumes, and even rubber.

The same is true of the tarry distillation products of coal. Originally tar was a useless and unwanted byproduct. If it was dumped on land it destroyed vegetation; if it was poured into rivers it killed the fish. Today this same coal tar provides the raw materials for many chemical industries.

←—— *A petroleum distillation unit is a massive intricate structure of stainless steel. The products of distillation are raw materials of mighty industries.* (Owen from Black Star)

16

OXYGEN AND HYDROGEN

In any discussion on burning one naturally thinks of oxygen and hydrogen. Hydrogen, either free or combined, is one of our commonest fuels. Oxygen is the ever-present substance that combines chemically with all fuels to cause a fire. Before we can understand the phenomenon of fire we must know something of the properties of these important elements. How can oxygen and hydrogen be prepared? Since water is a compound of these two elements, we might think of it as a suitable starting substance. Water, however, is a very stable compound; it does not decompose when heated, but merely changes to steam. So, having discarded water, we must look for convenient separate sources of oxygen and hydrogen.

Oxygen from the Air. A potential source of oxygen is, of course, the atmosphere which, you will recall, is a mixture composed chiefly of oxygen (21 per cent) and nitrogen (78 per cent). Our problem here would be to separate the oxygen from the nitrogen. This is not easy, although it is the way oxygen is prepared on a large scale in industry. In this process, the air is first liquefied by cooling it to a temperature of about −200° C. The boiling point of liquid nitrogen is −196° C and that of liquid oxygen is −183° C. Which substance escapes first if the temperature rises? Clearly the boiling point of nitrogen is reached if liquid air is heated through only a few degrees. At this temperature nitrogen boils off as a gas and liquid oxygen remains. Liquid oxygen can be stored for sometime in an open thermos flask. If the liquid warms up to −183° C, oxygen escapes as a gas. The gas is usually forced under pressure into steel cylinders where it is stored until needed.

Laboratory Preparation of Oxygen. To obtain oxygen from liquid air is far too expensive as a laboratory preparation. Instead, a white substance called potassium chlorate is used, since it readily decomposes, releasing oxygen when heated.

Demonstration

16-1. Oxygen from potassium chlorate

Place potassium chlorate, about half an inch deep, in a test tube. Clamp the test tube to a stand and then heat it gently. The potassium chlorate melts and then boils and gas bubbles escape. When this happens, hold a glowing wooden splint in the tube and notice that the escaping gas relights the splint. Now continue heating until no more gas is evolved. The residue,

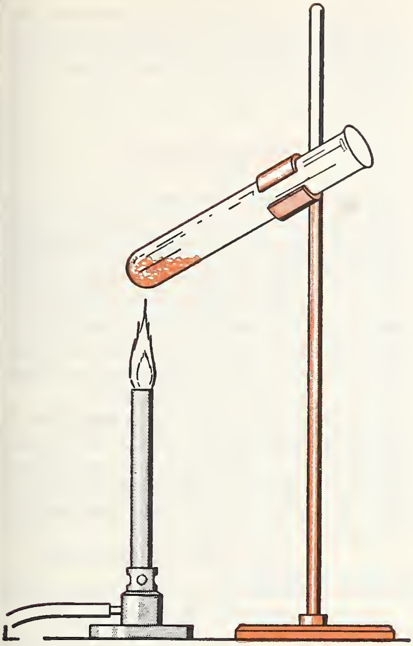


Fig. 16-1. Oxygen is released by heating potassium chlorate in a test tube.

potassium chloride, forms a white solid when the heating is stopped.

Several of the observations we have made need to be amplified. First, oxygen supports burning; indeed, a simple chemical test for oxygen is to see if the gas will make a glowing wooden splint burst into flame. The decomposition of potassium chlorate is a chemical reaction. This particular reaction can be represented by the word equation,

potassium chlorate →
potassium chloride + oxygen

Even endings in chemical words have meanings. Observe the endings ***-um, -ate,*** and ***-ide,*** the significance of which will be discussed in Chapter 18.

You probably noticed that to drive all the oxygen out of potassium chlorate it had to be heated to a very high temperature. This is wasteful of fuel gas and time. It is an advantage to decompose the substance at a lower temperature which can be done by adding ***a catalyst that alters the speed of a chemical reaction.***

Demonstration

16-2. The effect of a catalyst upon molten potassium chlorate

Again pour powdered potassium chlorate, about half an inch deep, into a test tube. Heat the tube cautiously until the potassium chlorate has just melted. Hold a glowing wooden splint in the test tube, and prove that no oxygen is being evolved. Now drop a little manganese dioxide into the molten chlorate. Observe the vigorous reaction and the evolution of a gas. Apply a glowing splint and notice that it instantly bursts into flame.

Manganese dioxide acts as a catalyst causing the potassium chlorate to decompose at a lower temperature, thereby speeding up the reaction. Catalytic reactions are, as we shall see later, very common and very important. Different substances are used as catalysts in different reactions. A catalyst is not permanently changed in a reaction, so that it can be recovered intact after the reaction has taken place. The best way to regard a catalyst is as a stimulant that speeds up reactions.

We can now prepare and collect oxygen, and observe the ease with which it combines with other elements. Since oxygen is a constituent of air, it is apparent that it is a colorless, odorless gas.

In chemistry we are frequently faced with the problem of collecting gases in bottles. This is done by first removing air from the bottle and then allowing the particular gas to enter. The usual procedure is first to fill the bottle with water, which, of course, excludes the air. A glass cover is placed on the bottle which is then inverted and lowered into a trough of water. If the cover is removed and the

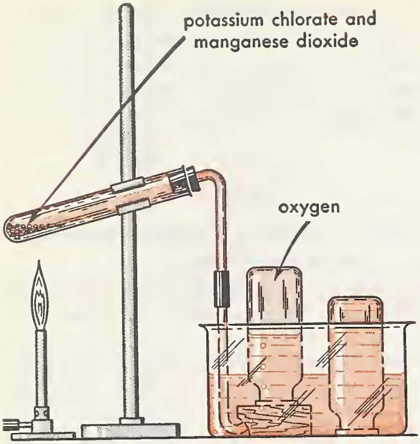


Fig. 16-2. Oxygen is collected in a bottle by displacing water.

bottle placed on a shelf, it is ready to receive the gas through a glass delivery tube. The arrangement is shown in Fig. 16-2.

Demonstration

16-3. To collect oxygen

Pour a mixture of potassium chlorate and manganese dioxide, about an inch deep, into a test tube and arrange the apparatus as shown in Fig. 16-2. Heat gently and gas bubbles are collected at once. When the bottle is filled with gas, replace it by another containing water. This first bottle contains some air from the test tube and should be discarded. When the second bottle is filled with oxygen, slide on the glass plate while it is still underwater, and lift the bottle out of water, invert it, keeping it covered. Collect three bottles of oxygen in this way.

Oxygen Supports Combustion. The most important property of oxygen is that it supports burning or combustion. The simplest combustion reactions are those in which elements combine with oxygen. Whenever an element combines with oxygen, an ***oxide*** is formed. Or, we can say, ***an oxide is a compound of oxygen and one other element.*** The three elements we shall select are magnesium, phosphorus, and carbon.

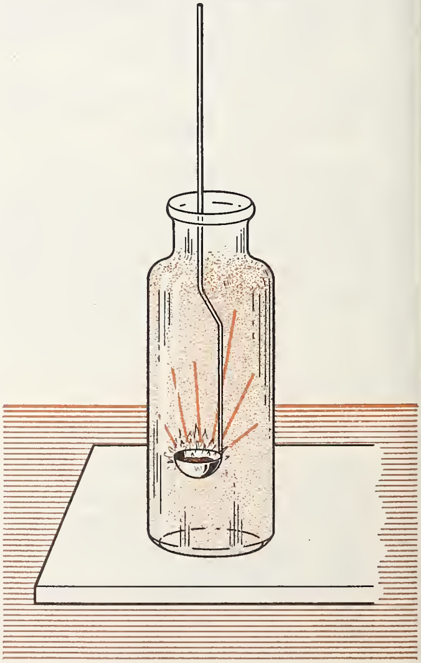
Demonstration

16-4. The combustion of elements

Phosphorus. Place a little red phosphorus in a combustion spoon and lower it into a bottle of oxygen. Nothing happens. Now heat the phosphorus till it just begins to burn and then lower it into the oxygen. The burning is speeded up as shown by the intense white flame. Notice that a white powder is formed in this reaction.

Now repeat the experiment using first a few short strips of magnesium ribbon and then a piece of charcoal. The magnesium burns in oxygen with a blinding flame forming a white powder; the hot charcoal, on the other hand, merely glows.

Fig. 16-3. Phosphorus, in a combustion spoon, burns in a bottle of oxygen.



Notice that, as a rule, most elements do not react rapidly with oxygen at ordinary temperatures. Or, stated another way, oxygen is relatively inactive at ordinary temperatures. Why is this? We shall say more on this subject in the next chapter. For the time being, we will merely state that oxygen molecules show little chemical activity at ordinary temperatures. If oxygen is heated, however, some molecules are changed to atoms which are exceedingly active.

The white powder formed from phosphorus is phosphoric oxide and the equation for the reaction is

phosphorus + oxygen →
phosphoric oxide

The corresponding reactions with magnesium and charcoal are

magnesium + oxygen →
magnesium oxide

carbon + oxygen → carbon dioxide

Oxidation. Under the right conditions oxygen combines with most of the elements and with many compounds. All these are called ***oxidation*** reactions. If the oxidation is slow, as in the rusting of iron exposed to damp air, the heat of the reaction is slowly dissipated into the air and no flame is produced. If the oxidation is rapid, the temperature rises so fast that the combustible substance is consumed in fire. In other words, combustion, or burning, is an indication of rapid oxidation.

What Is Oxygen Used For? Oxygen is used to produce high temperatures. For instance, oxygen combines with acetylene very vigorously. In this reaction the oxyacetylene flame reaches a temperature of almost 3500° C, one of the highest temperatures that can be produced by chemical means. The oxyacetylene flame is used for cutting and also welding steel, an operation in the construction of ships and skyscrapers.

Oxygen is the breath of life. The oxygen we inhale passes through the lung membrane and enters the blood stream. It is carried to every part of the body where it unites with the food materials in the cells, forming carbon dioxide and water, and releasing heat. If the lung is damaged or its capacity reduced by fluid, as in pneumonia, atmospheric oxygen cannot pass into the blood at the proper rate. The patient is then placed in an oxygen tent.

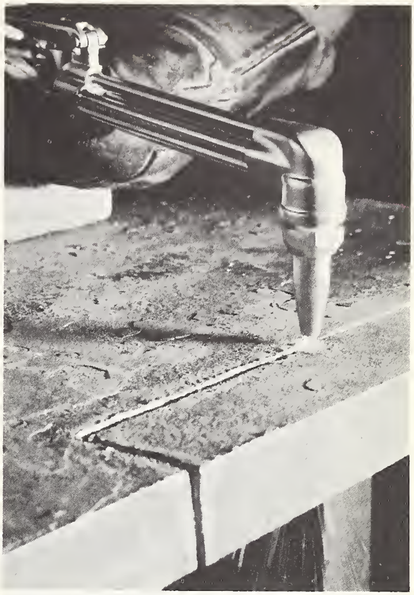


Fig. 16-4. An oxyacetylene flame easily cuts through a 2-inch steel plate.

Hydrogen. Although hydrogen is free and abundant in the sun and stars, there is no free hydrogen on earth. As stated in Chapter 3, it is probable that hydrogen was present in the earth's original atmosphere, but because the atoms are so light the earth's gravitational pull was unable to hold them and, in consequence, they escaped into space. One might ask, *Why* is hydrogen the most abundant element in the universe? Atoms are made up of fundamental particles, protons, neutrons and electrons. Hydrogen atoms are the simplest and the lightest of all atoms; they each consist of a proton and an electron. Hydrogen atoms are therefore the most fundamental particles and, as a result, the most abundant.

Although hydrogen is not free like oxygen, it occurs in a vast number of compounds. It occurs in all organic compounds such as sugars, starches, fats and proteins, in fossil fuels such as coal and oil, and in many inorganic compounds such as water and acids. In industry, where it is prepared on a large scale, hydrogen is obtained from water; in the laboratory it is obtained more easily from an acid.

What Is an Acid? The first acid used by man was probably acetic acid which occurs in vinegar. Vinegar is formed when wine oxidizes and turns sour. It was known to the Greeks and Romans; indeed, it was an ingredient of one of their favorite drinks.

Hundreds of acids are known, many made by natural processes. Most fruits contain acids which accounts for their sharp flavor. Apples contain malic acid; lemons, citric acid; gooseberries, tartaric acid. As stated on page 59, acids are also formed by the decomposition of vegetation. If decomposition is excessive, the soil may become sour and lose its fertility. This is particularly true of acid soils formed from pine needles.

In industry the most important acids are sulfuric, hydrochloric, and nitric acids. These acids are manufactured from minerals and, for this reason, are called ***mineral acids.*** They may be ***concentrated,*** in which case a relatively small proportion of water is present with the acid, or ***dilute,*** in which case a large proportion of water is mixed with the acid.

To Prepare Hydrogen. The most convenient source of hydrogen in the laboratory is dilute sulfuric acid. The hydrogen can readily be displaced by a suitable metal such as zinc. This method of preparing hydrogen is suggested by the definition of an acid, which should be memorized. ***An acid is a compound which contains hydrogen, this hydrogen being replaceable by a metal to form a salt.***

If dilute sulfuric acid is poured onto a piece of zinc in a test tube, a vigorous effervescence takes place and bubbles of hydrogen escape.

A word equation for the preparation of hydrogen is

$$\text{zinc} + \text{sulfuric acid} \rightarrow \text{zinc sulfate} + \text{hydrogen}$$

The compound zinc sulfate is a ***salt,*** a substance referred to on an earlier page. If we wish to collect a bottle of hydrogen we must proceed as follows:

Demonstration

16-5. To collect hydrogen

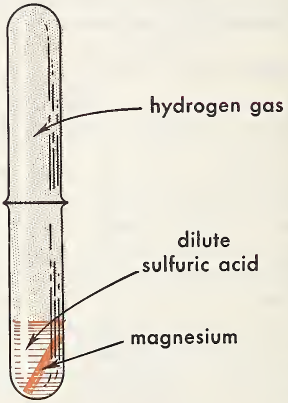


Fig. 16-5. The preparation of hydrogen.

Fill a test tube about one-third full with dilute sulfuric acid and drop a strip of magnesium ribbon, one to one and a half inches long, into it. In doing this, hold the test tube in a vertical position so that the strip of magnesium drops straight into the acid. If the test tube is held at an angle, the magnesium will stick to the side of it where no reaction can take place. If this happens, shake the test tube gently to wash the magnesium down. Be very careful not to splash any of the acid out of the test tube.

As soon as the magnesium comes into contact with the dilute sulfuric acid, a violent

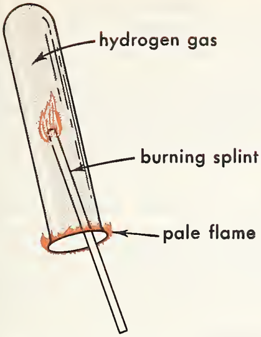


Fig. 16-6. The safety test for hydrogen.

chemical reaction takes place and hydrogen gas is given off. This reaction is

magnesium + sulfuric acid →
magnesium sulfate + hydrogen

The reaction takes place for a few seconds only. Therefore, to collect the hydrogen given off, make sure you have another test tube handy and place it mouth to mouth over the first test tube as shown in Fig. 16-5.

Hold the test tube of collected hydrogen with its mouth downward and insert a burning splint into it. Note that the flame goes out, but you may see hydrogen burning with a pale blue flame at the mouth of the test tube as shown in Fig. 16-6.

Now turn the test tube of hydrogen to an upright position for five to ten seconds. Insert a burning splint again. A loud "bark" indicates that air is mixed with the hydrogen. (In doing this experiment, make sure that you do not point the test tube towards your face or that of your partner.) The reaction for the burning of hydrogen in oxygen is

hydrogen + oxygen →
water vapor + heat

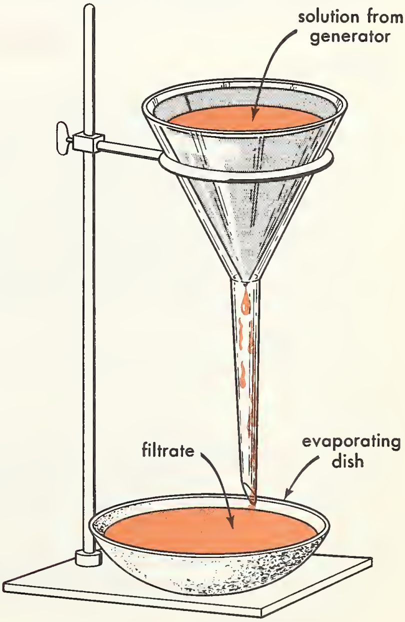
The above experiment shows that a mixture of air and hydrogen has a dangerous explosive force in a closed container. However, the burning splint is extinguished in pure hydrogen. From this we conclude that ***hydrogen does not support burning***. Finally, ***hydrogen*** burns. Notice, however, that the hydrogen flame is at the mouth of the test tube. How can we explain this? Clearly by saying that hydrogen needs oxygen in order to burn.

We can now compare hydrogen and oxygen as follows: ***Hydrogen burns but does not support burning; oxygen does not burn, but does support burning.***

Filtration. Repeat the above experiment for the preparation of hydrogen, but this time drop some pieces of zinc into dilute sulfuric acid. Note that the chemical reaction is less vigorous than it was with magnesium. This is because zinc is a less active metal than magnesium. Look up an activity table and note the positions of the two metals on it.

Note that the solution in the test tube in which the zinc and acid were mixed contains a good deal of solid material that can be filtered out by pouring it into a cone of folded filter paper supported in a glass funnel. Between the fibers of the filter paper there are spaces small enough to hold back tiny particles of solid yet large enough to allow a liquid to pass through. Removing suspended particles from a solution is known as ***filtration***, and

Fig. 16-7. Filtration.



the clear filtered solution is called the ***filtrate.***

Demonstration

16-6. To prepare zinc sulfate crystals

Filter the solution in the test tube. Catch the filtrate in an evaporating dish as shown in Fig. 16-7. Put the filtrate aside for a day or so to evaporate and then observe the needle-shaped crystals of zinc sulfate that appear in the dish.

Why have crystals formed? The water evaporates on standing so that the solution becomes more and more concentrated; that is, the solution becomes richer in zinc sulfate as evaporation proceeds. Eventually the solution contains as much zinc sulfate as it can hold. It is then ***saturated.*** If evaporation proceeds beyond the saturation point crystals begin to form, and as evaporation continues the crystals grow in size.

Things to Remember

A ***catalyst*** is a substance that alters the speed of a reaction.

An ***oxide*** is a compound of oxygen and another element.

In ***rapid oxidation*** a flame is produced.

An ***acid*** is a compound which contains hydrogen, this hydrogen being replaceable by a metal to form a salt.

Hydrogen burns, but does not support burning; ***oxygen*** does not burn, but does support burning.

Filtration is the removing of solid particles from a solution.

Oxygen is prepared by decomposing potassium chlorate.

Hydrogen is prepared by decomposing an acid by a metal.

Questions

GROUP A

1. What is the essential element in (*a*) an acid, (*b*) an oxide?
2. Define each of the following terms, (*a*) acid, (*b*) oxide.
3. The temperatures concerned with liquid air are −183° C, −196° C, and −200° C. What do each of these temperatures signify?
4. Write word equations for the following reactions: (*a*) preparation of oxygen (*b*) preparation of hydrogen (*c*) carbon + oxygen (*d*) the hydrogen flame.
5. What is a catalyst? Give an example.
6. Name two organic compounds and two inorganic compounds that contain hydrogen.
7. Name two mineral acids and two acids that are not mineral acids.

GROUP B

8. Draw and label a diagram to show how you would prepare oxygen.
9. Concerning your diagram in Question 8, explain (*1*) why water does not fall out of the inverted bottle, and (2) why gas bubbles are able to replace this water.
10. What is the most characteristic property of oxygen? Illustrate your answer by referring to a reaction that takes place inside the human body and one that does not.

11. (*a*) Draw and label a diagram for the preparation of hydrogen. (*b*) Where does the hydrogen come from, the acid or the zinc? Give a reason for your answer.

12. Compare and contrast the characteristic properties of oxygen and hydrogen.

13. (*a*) How are crystals of zinc sulfate prepared? (*b*) Compare this method of crystal formation with the formation of mineral crystals as in granite.

Things to Do

A test for oxygen • Prepare a water solution of pyrogallic acid and potassium hydroxide. This solution will be found to be colorless. Pour it into a jar of oxygen and shake the jar. Note that the oxygen dissolves readily in the solution, and at the same time turns this colorless liquid to a deep brown color. No other gas will behave in a similar way. Try it with carbon dioxide and hydrogen.

Since oxygen is the only gas that will turn a colorless solution of pyrogallic acid and potassium hydroxide to a deep brown color, we may consider this as a conclusive test for oxygen.

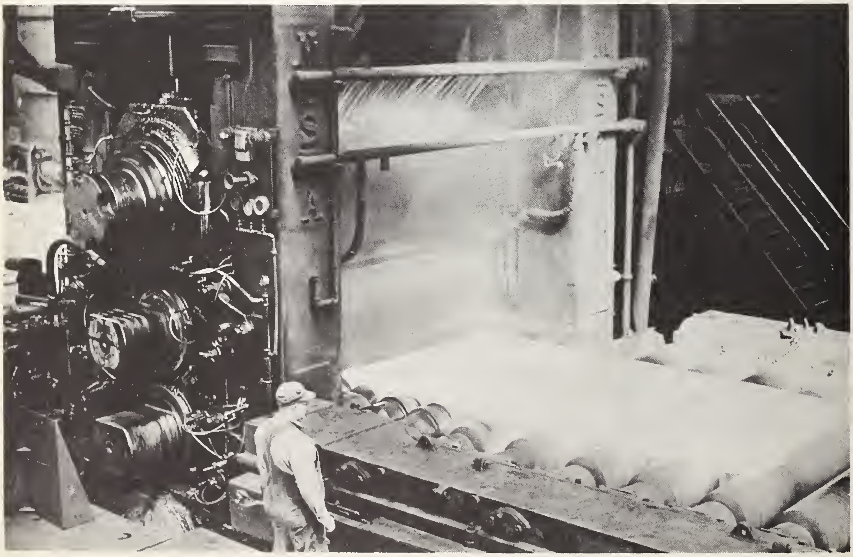
The above experiment may also be done by bubbling gas through the solution.

17

WATER, AN IMPORTANT COMPOUND

Have you ever wondered why a river flows through every sizable city? The sites of most cities were originally selected because of their nearness to rivers. Why is this? Industries concentrate in cities, and industries need water, vast quantities of it. They need it as a cleanser, as a solvent, as a coolant, and to carry away wastes. Such industries as papermaking, sugar refining, soap making, textile manufacture, the refining of metals and the manufacture of fuel gas all need it for one or

Fig. 17-1. The steel industry uses more than five billion gallons of water a day. This picture shows a white-hot slab being rolled to a plate about an inch thick. The jets of water wash off any scale that forms on the plate and also control the rate of cooling.



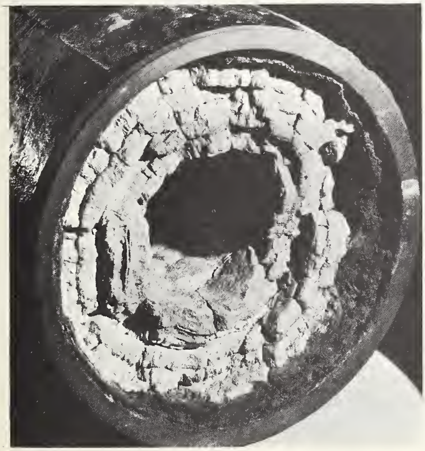
more of these purposes. Industrial plants in the United States need almost 70 billion gallons of water per day. Even an average-sized North American coke plant uses about 10 million gallons of water a day, enough to supply a city of 100,000 people.

Water used in industry must be selected with care. Some rivers, like the Mississippi, carry a great deal of sediment. The waters of others may be hard. Hard water carries certain minerals in solution that make it difficult to form a lather with soap. Both sediment and hardness are objectionable features. Sediment can, of course, be removed by filtration and hardness can be removed by treatment with chemicals. But these are expensive processes and, whenever possible, it is better to select water that is soft and fairly free from sediment.

Hard and Soft Water. ***Hardness in water is caused by compounds of calcium or magnesium in solution.*** The water in limestone (calcium carbonate) regions, for example, is usually very hard.

When soap is shaken with hard water a scum or curdy precipitate is formed. This is because the soap reacts with the calcium or magnesium compounds and, as a result, it is unable to do its proper work of cleansing. The formation of scum therefore wastes soap and, in industrial plants, this wastage can be very costly.

Fig. 17-2. Boiler scale in a 6-inch pipe. The scale is composed largely of calcium carbonate deposited from hot hard water that flows through the tube.



Another objection is that when hard water is heated minerals are deposited. For example, calcium carbonate, from hard water, often forms a lining or hard scale in the tubes of boilers. This has two serious effects. First, the scale is a poor conductor of heat, so that only a fraction of the heat of the fire passes through to the water in the tubes. This entails large heat losses and a waste of fuel. It is said that a scale only one-quarter of an inch thick increases fuel consumption by almost 50 per cent. Then, again, the scale impedes the flow of water through the tubes and this might cause an explosion in the boiler.

Water can be softened by adding washing soda (sodium carbonate) to it. The sodium carbonate forms an insoluble compound with the dissolved minerals so that hardness is removed. For some types of hardness, lime, which is much cheaper than washing soda, is commonly used.

Demonstration

17-1. To harden and soften water

Almost fill a test tube with distilled water and then add a few drops of calcium chloride solution to harden the water. Pour half of the water into another test tube. Half fill a third test tube with distilled water and place all three test tubes in a rack.

Pour a little soap solution into one of the tubes containing hard water and also into the tube containing distilled water. Shake the tubes. In the distilled water a lather is formed; in the hard water, a scum is formed with little lather.

Now add a little sodium carbonate solution to the remaining tube of hard water. The white precipitate that forms contains the mineral causing hardness. Filter the water, add a little soap solution to the filtrate, and shake. A lather is now formed. Can you now explain why the addition of washing soda softens hard water?

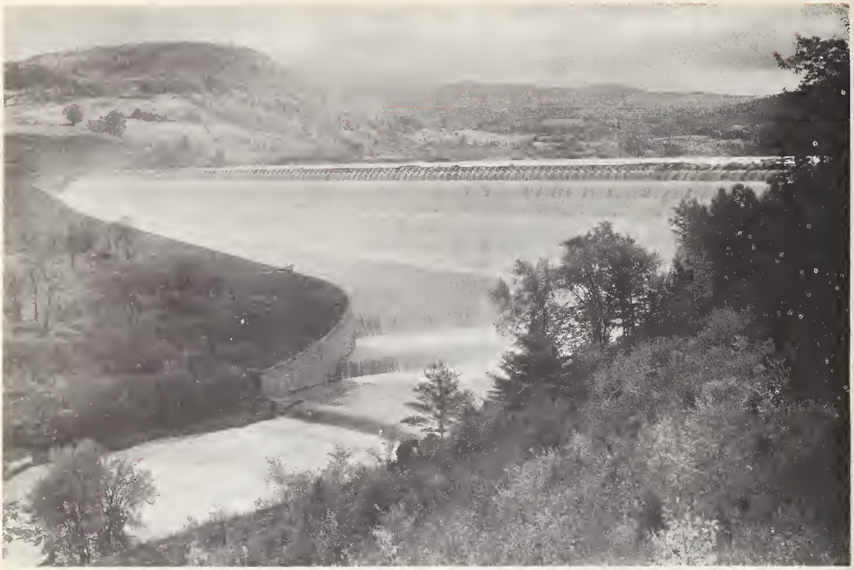


Fig. 17-3. Flood water discharging over the spillway of Schoharie Reservoir in the Catskills, part of the New York water supply system. Notice the sparsely populated water shed in the background.

Drinking Water. As human beings, we need large quantities of drinking water. Most of it comes from rain that falls on the land, seeps through the soil, and eventually joins a stream, river, or lake. In seeping through soil it may become contaminated.

Soil contains a vast number of bacteria —some helpful, some harmful. Typical harmful bacteria are typhoid and dysentery germs. Both can live in water. If this water is drunk, the bacteria enter our intestines where they thrive and multiply. This may cause typhoid fever, a dangerous and often fatal disease. Fifty years ago there were numerous cases of typhoid but, as a result of the sterilizing of drinking water, the disease has been almost wiped out.

Supplying drinking water to a large city is a major operation that concerns engineers, chemists, and bacteriologists. An interruption in the water supply of a city could lead to calamity. Factories would be closed, fires could not be controlled, and lack of sanitation might cause widespread disease. Drinking water must be available in vast quantities. The amount provided for each person should be at least 100 gallons per day; New York City, for example, uses almost one billion (1,000,000,000) gallons of drinking water every day.

Drinking water for cities is collected on watersheds—large forested areas in sparsely populated regions. Every precaution is taken to see that the water is not polluted by sewage, a common source of typhoid bacteria. The water drains from watersheds and is stored in reservoirs which are artificial lakes formed by constructing a dam at the lower end. During the spring rains water accumulates and, usually, enough is collected to meet needs throughout the dry season. A reservoir also helps to purify the water since suspended organic matter slowly settles in its quiet waters. From the reservoir the water is carried, usually by underground aqueducts, to the city filtration plant, where the purification process takes place.

The Purification of Drinking Water. The purification process depends upon the nature of the raw water. After purification water should be clear, soft, tasteless, odorless and free from harmful bacteria. Purification may involve any or all of the following treatments: (1) sedimentation, (2) filtration, (3) aeration, (4) lime-soda treatment, (5) chlorination.

Sedimentation. In this process raw water is run into concrete tanks about 15 feet in depth. Aluminum sulfate, a white powder, is then added to the water and reacts with it to form a jelly-like mass. This gelatinous precipitate slowly settles, carrying most of the sediment and bacteria to the bottom of the tank. About once a month the tank is thoroughly cleaned to remove the deposit that collects.

Demonstration

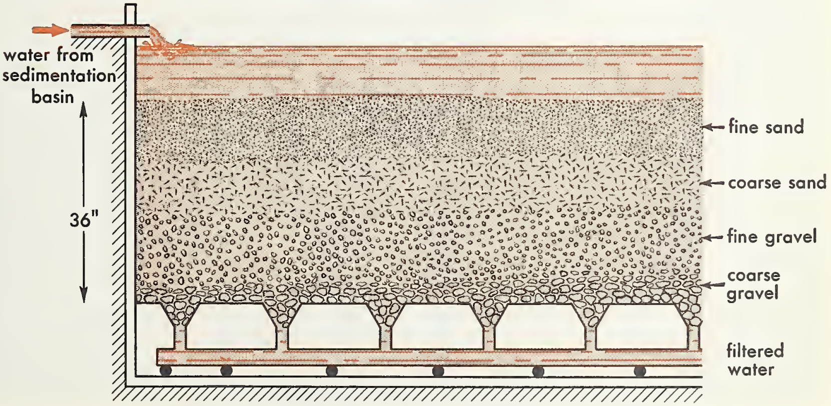
17-2. Sedimentation

Almost fill two glass cylinders (hydrometer jars) with water. Add a little ammonium hydroxide to each jar so that the water is slightly alkaline to litmus (that is, until the water turns red litmus paper blue). Add some dirt, fine sand, and powdered clay (kaolin) to both jars, and stir. Pour about 20 cc. of aluminum sulfate solution into one of the jars. After standing for some time, preferably overnight, notice that, in the jar containing aluminum sulfate, the sediment has been carried to the bottom. In the other jar, however, much of the sediment still remains in suspension.

Filtration. After standing in the sedimentation basin for two or three hours the water is run to the filter plant, which consists of a number of concrete basins similar to that shown in Fig. 17-4. Each basin is packed with graded gravel and sand to a depth of about three feet—coarse gravel at the bottom, then fine gravel, coarse sand, and finally fine sand at the top. Water, which runs in at the top, is so effectively filtered that, when it emerges at the bottom, it is free of all sediment and most of the remaining bacteria. The filters are usually cleaned every day. To do this, clean water is forced in at the bottom and up through the sand. The wash water escapes at the top and is rejected.

Aeration. The purpose of aeration is to remove odors and tastes by oxidizing the

Fig. 17-4. One of the processes in the preparation of drinking water is the filtration of raw water through sand and gravel.



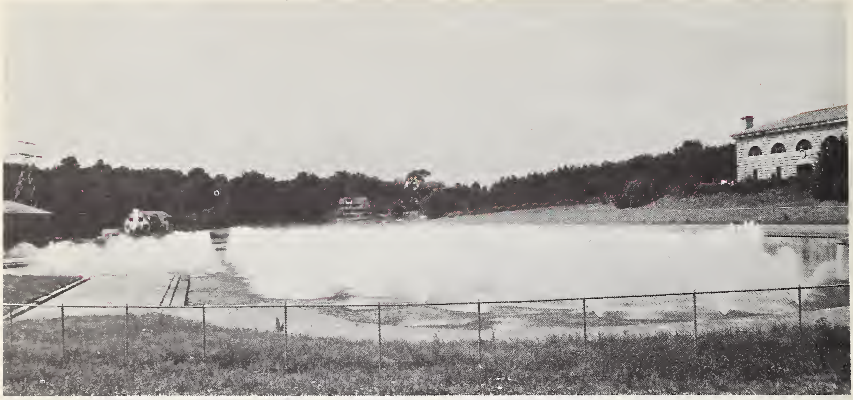


Fig. 17-5. The Kensico Aerator on the New York City water system. Water is thrown into the air through jets to oxidize organic impurities.

offending organic material. This is done by cascading the water over an artificial waterfall or throwing it high into the air through jets. In this way organic material is exposed to air and sunlight, and oxidized.

Lime-Soda Treatment. If analysis shows that the water is too hard for domestic or industrial use it is softened by the addition of lime (calcium hydroxide) and soda (sodium carbonate) which react with the calcium and magnesium compounds to form a precipitate. This precipitate is then removed by filtration.

Chlorination. Chlorination, which consists of bubbling chlorine gas through the filtered water, is the final operation in the purification process. Chlorine is a poisonous gas so that only a small quantity is used—enough to destroy any remaining harmful bacteria, but not enough to be harmful to human beings. After this treatment, the water is free from contamination and fit to drink.

The term ***purification*** is somewhat misleading in connection with this process. Really pure water contains no substances in solution; drinking water on the other hand contains a good deal of dissolved material. Indeed, the substances in solution add to the flavor and make water palatable. How can we obtain pure water? Or, how can we remove the dissolved solids from ordinary water? This is done by a process called ***distillation.***

How Water Is Distilled. If a solution is boiled, water vapor escapes and the solid impurities are left behind. If the vapor is sent through a condenser (which is a tube surrounded by a cold-water jacket) the vapor is condensed back to water. This is distilled water. Distillation thus consists of two operations (1) boiling the solution and (2) condensing the vapor, as is shown in the following experiment.

Demonstration

17-3. Distillation of a salt solution

The distillation apparatus consists of three parts: a boiler or distilling flask, a condenser, and a receiver. Half fill the flask with salt solution, and set up the apparatus as shown in Fig. 17-6. Boil the solution and collect some of the condensed vapor. Taste the distilled water. It has no flavor because it contains no dissolved solids.

Distilled water is used in the manufacture of some chemicals, in making up medicines, and in automobile batteries. It is also used in the boilers of high pressure steam plants, where a deposit of

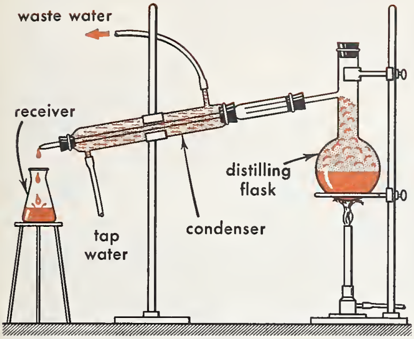


Fig. 17-6. Distillation consists of boiling a solution and condensing the vapor.

scale must be avoided. But, for most purposes, this high degree of purity is unnecessary.

Kinds of Water in Nature. Natural waters can be conveniently divided into three groups: (1) rain water, (2) surface waters, and (3) sea water.

Rain Water. Rain water is the purest of natural waters; it is simply condensed water vapor. However, it is difficult to collect and it is not a convenient source of drinking water. But in places like Bermuda where there is a shortage of surface water, the collecting of drinking water is a local problem. Here the roofs of houses are flat, and painted white, and generally designed to catch clean rain water.

Surface Waters. Surface waters, which include streams, rivers, and lakes, are by far the most important source of water for all purposes. The amount of solid in solution depends upon the solubility of the rocks over which the streams flow. If a river flows over an almost insoluble bed like granite, it is relatively free from dissolved impurities. As a rule, however, river water contains a large number of compounds in solution. Of these, calcium compounds are the most common. Can you explain why?

Fig. 17-7. Surface water. Cameron Falls cascading over cliffs in Waterton National Park, Alberta.

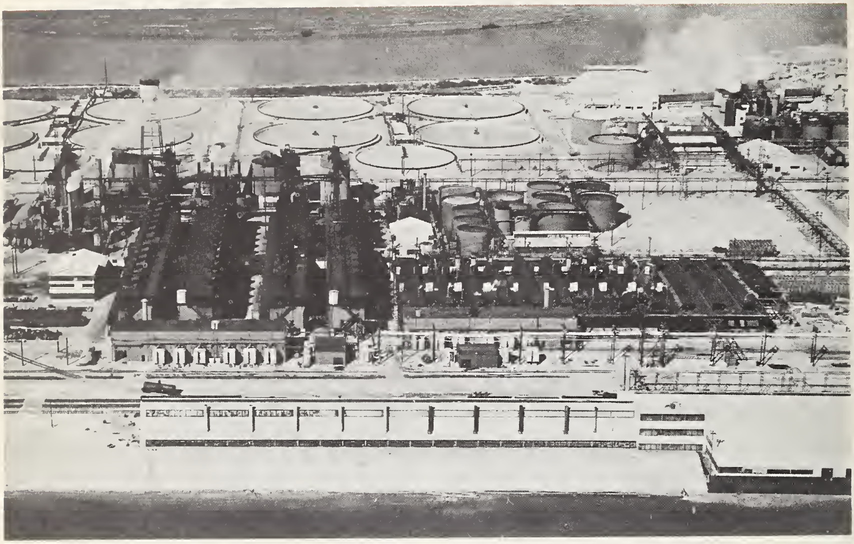


Fig. 17-8. Magnesium from sea water. The circular tanks in the background contain sea water from which the metal magnesium is obtained.

Sea Water. The sea is the reservoir for all dissolved solids carried down by rivers, and its disagreeable taste is due to these substances in solution. Elements that form the most plentiful compounds in sea water are sodium, potassium, magnesium, calcium, chlorine. By far the most abundant metallic element is sodium, present as sodium chloride. However, the most abundant element in surface waters is calcium derived, you will recall, from the erosion of limestones and certain granites. Here then is a paradox. Calcium is the most plentiful element in river water, whereas sodium is the most plentiful in sea water. Can you explain the paradox? The explanation is, of course, that in the sea many kinds of small animals build hard shells of calcium carbonate from the calcium compounds in solution. Indeed, geologists have computed that, in this way, almost one and a half million tons of calcium are removed from the sea every year.

The Sea as a Storehouse. The sea is a storehouse for everything washed from the land. Consequently, sea water contains practically every known element in one compound or another; even gold and uranium are present in sea water in relatively small amounts. The amounts of six of the most plentiful elements present, in the form of compounds, in one cubic mile (4,700,000,000 tons) of sea water is shown in the following table:

Chlorine	62,000,000 tons
Sodium	40,000,000 tons
Magnesium	5,500,000 tons
Calcium	1,700,000 tons
Potassium	1,600,000 tons
Bromine	270,000 tons

These are huge quantities, and sea water appears to be a promising source of raw materials. Can these elements be recovered from sea water and put to use? This is an important question since the mineral resources in the earth are being rapidly used up. Minerals cannot be renewed like things that grow, like trees for instance. Once minerals are mined, the earth is, to that extent, the poorer. Already there is a world shortage of copper, lead, tin, and tungsten, and some day engineers will have to turn to the sea for the metals that are needed.

So far only two elements, bromine and magnesium have been profitably extracted from the sea water. Magnesium is used as an alloy particularly in the construction of planes, and bromine is needed in the antiknock liquid added to gasoline. The great difficulty in the large scale extraction of elements from sea water is that they are thinly scattered. For example, one ton of sea water yields only 2.5 pounds of magnesium and only one ounce of bromine. Extraction is therefore an expensive operation. Powerful pumps are needed to force vast quantities of sea water through the factory which at best gives a poor yield of product. Nevertheless, as metals become scarcer, methods for extracting metals from sea water will be able to compete with present industrial processes.

Salt, a necessary substance in the life process, has been obtained by evaporating sea water from earliest times. Man needs about two-thirds of an ounce of salt every day; without it he will die. Early man learned by experience that salt is necessary to life. It came to be regarded as a gift of God, and much superstition was associated with it. Salt is steeped in the history of all races. It was an article of barter between tribes: it was a medium of payment; indeed, the word *salary* is derived from the word *salt*. Salt has been the cause of bloodshed and even war; one of the causes of the French Revolution was a tax on salt.

Today salt is needed in many industries. In most countries, however, it is easier to mine salt from underground deposits than to evaporate sea water. Salt deposits are the remains of inland seas that dried up ages ago. In Canada salt deposits are found in many provinces, particularly in Alberta, Ontario, and Nova Scotia, where it is recovered on a large scale.

Water, a Natural Resource. It is apparent that the health of a community, its development and its prosperity depend on an abundance of clean water. The increase in population, industrial expansion (with its ever-increasing demand for water) and the need for irrigation have forced us to realize how dependent we are upon water. Water is one of the nation's most important natural resources, and its conservation is now regarded as a national problem. In many parts of North America the water table is dropping. In New England, where water abounds, the water table has dropped five feet or so in the past 10 years; in Arizona, one of the driest states, it has dropped 50 feet in the same length of time.

The best way to conserve water is to preserve the watersheds. Watersheds are vast upland areas covered by shrub and trees that collect rain water and insure a steady and abundant supply of water. If watersheds are damaged by fire, by excessive cutting of timber, or by overgrazing, our water supplies will be impaired.

Even in our homes a great deal can be done to prevent waste of water. One wasteful practice results from assuming that taps are made to be turned on indiscriminately. We all know that, particularly in hot weather, drinking water is allowed to run down the drain until it is "good and cold." Only then is it used for drinking purposes. Millions of gallons of drinking water could be saved every summer by keeping a bottle of water in the refrigerator. This is a simple precaution worth putting into practice when reservoirs are low.

Things to Remember

Hardness in water is caused by calcium and magnesium compounds. Hardness can be removed by washing soda or lime.

To change **raw water to drinking water** the following steps are usually necessary: (1) sedimentation, (2) filtration, (3) aeration, (4) lime-soda treatment, (5) chlorination.

The purification of water by **distillation** involves boiling the solution and condensing the vapor.

Natural waters may be classified as (1) rain water, (2) surface water, (3) sea water.

Calcium is the most abundant element in surface waters; **sodium** the most abundant in sea water.

Questions

GROUP A

1. What is (*a*) hard water, (*b*) soft water?
2. What is the most effective way of destroying harmful bacteria in drinking water?
3. What is the purpose of (*a*) alum and (*b*) aeration in the purification process?
4. What is meant by the terms (*a*) **distillation** (*b*) **surface waters?**
5. Name three uses for distilled water.

GROUP B

6. (*a*) What is scale in boiler tubes? (*b*) Why is it objectionable? (*c*) How can it be prevented?
7. (*a*) How is hard water softened? (*b*) What happens in the softening process? (*c*) How would you test water to see if it is soft?
8. Draw a distillation set-up and explain how it operates.
9. Explain why calcium is usually the most abundant element in river water but not the most abundant element in sea water.
10. Name the two substances (other than sodium chloride) that are obtained from sea water. Discuss some of the reasons why extraction from sea water is an expensive process.
11. Why is water conservation important? Discuss some of the methods of conserving water.

Things to Do

Inspecting a water purification plant. Visit your local municipal water purification plant. Ask where the watershed is. Is the raw water piped from the watershed to the plant? Is there a reservoir for water storage?

Examine all the purification processes, particularly chlorination. Notice how the drinking water is stored and protected before it is piped to the city.

Finally, if there is a nearby standpipe observe how drinking water is pumped into it. And notice also that the water level in the standpipe is higher than all the faucets in town. Why?

Make a written report of the purification process in the plant. Draw diagrams where they can be helpful in clarifying your report.

18

A CHEMIST'S LANGUAGE AND SHORTHAND

IN CHEMICAL REACTIONS compounds may be decomposed or new compounds may be formed; atoms, however, are never decomposed—they merely exchange places. The activity of atoms is the basis of all reactions, and literally thousands of compounds can be made from relatively few atoms, only a hundred or so. The chemist has given special attention to these "building blocks" that make up compounds. He has invented a shorthand of signs and symbols, and, once their meaning is understood, chemical reactions can be expressed in a simple, meaningful, symbolic language.

You may have already noticed that, in chemical language, word endings are significant. For instance, the ending ***-um*** usually indicates a metal; aluminum, magnesium, and sodium are metals. The ending ***-ide*** is used in naming compounds. It means that a compound with this ending is made up of ***two,*** and only two elements. The compound sodium chloride, for example, is made up of the elements sodium and chlorine; carbon dioxide is a compound of carbon and oxygen. Still another common ending for a compound is ***-ate.*** It means that this particular compound is made up of ***three*** elements, one of which is oxygen. Zinc sulfate, for instance, is a compound made up of the elements zinc, sulfur and oxygen, and it should be apparent that the elements present in calcium carbonate are calcium, carbon and oxygen.

Symbols. The chemist also uses symbols to represent elements. Wherever possible he uses the first letter of the name, capitalized, as the symbol. For instance, the symbol for hydrogen is H; for oxygen, O; for nitrogen, N; carbon, C; phosphorus, P; and sulfur, S.

If more than one element begins with the same letter, a second ***small*** letter may be added to avoid confusion. Thus, the symbol for calcium is Ca, and for chlorine, Cl. Other common elements with two-letter symbols are aluminum, Al; magnesium, Mg; zinc, Zn; manganese, Mn; radium, Ra; bromine, Br.

The symbols of a few elements are derived from their corresponding Latin names. These include iron, Fe (ferrum); copper, Cu (cuprum); lead, Pb (plumbum); silver, Ag (argentum); sodium, Na (natrium); mercury, Hg (hydrargyrum); potassium, K (kalium).

Valence. It has already been stated that the formula for water is H_2O. Why not HO or HO_2? Atoms of a given element have a certain limited combining power which can be represented by bonds. These bonds are electrical in nature. A hydrogen atom has one bond and can be written H—; an oxygen atom has two bonds and can be written —O— or O=. A substance with free bonds is chemically active. Hydrogen atoms are therefore active; that is why they combine with each other to form a stable molecule, H—H, or H_2. Similarly, oxygen atoms are active and combine to form a stable molecule, O=O, or O_2. However, if hydrogen and oxygen atoms are intermixed they prefer to combine with each other rather than with themselves. In this case the compound H—O—H is formed. It can be written HOH or H_2O and, as a result, water is a stable compound.

It should now be apparent that before we can write the formulas of compounds we must know the number of bonds in the atoms of the constituent elements. The number of bonds of an atom is usually called its ***valence.***

Or, we can say, ***the valence of an element is the number of bonds by which one atom of this element can attach itself to other atoms.***

The valences of elements vary. Thus the valence of hydrogen is one, of oxygen, two, of aluminum three, of carbon four, and of phosphorus five. The following is a list of the valences of some of the common elements:

Valence	*Element*
One	Hydrogen (H), sodium (Na), potassium (K), chlorine (Cl), bromine (Br)
Two	Oxygen (O), sulfur (S), copper (Cu), magnesium (Mg), zinc (Zn), calcium (Ca)
Three	Aluminum (Al), iron (Fe)
Four	Carbon (C)
Five	Phosphorus (P)

We can now write formulas for some of the common compounds.

Sodium chloride. The valence of sodium is one, and chlorine is one. Hence, one sodium atom combines with one chlorine atom, and therefore the formula is NaCl.

Calcium chloride. The valence of calcium is two, and so one calcium atom will combine with two chlorine atoms, giving the formula $CaCl_2$.

Magnesium sulfide. Magnesium and sulfur both have a valence of two, so here the formula is MgS.

Radicals. In all the above examples the compounds contain only two elements. Sometimes, however, compounds contain three or four, or even more elements. Bonds between some kinds of atoms are strong and between others much weaker. Strong bonds are difficult to break even in chemical reactions. A group of elements which form part of a compound and are linked together by strong bonds is called a ***radical.*** Or, we can say, ***a radical is a group of two or more elements that usually behaves like a single element during a chemical reaction.*** That is, radicals are not usually decomposed in chemical reactions.

A common radical is the sulfate radical. Its formula is SO_4 and its valence is 2. It is important to remember that radicals are parts of compounds; they cannot exist by themselves. For example, there is no such substance as sulfate. There is, however, a substance called hydrogen sulfate. What is the formula for hydrogen sulfate? The constituent parts of this compound are H of valence one and sulfate of valence two. Therefore its formula is H_2SO_4. You may recall that the common name for hydrogen sulfate is sulfuric acid.

Similarly the formula for zinc sulfate is $ZnSO_4$. What is the formula for aluminum nitrate? The valence of aluminum is 3 and nitrate, NO_3, is 1. The smallest number of valence bonds needed to bind these parts together is 3. Hence 1 atom of aluminum and 3 nitrate radicals are needed. Therefore the formula is $Al(NO_3)_3$. Notice that since the subscript three applies only to the nitrate radical it is necessary to enclose this radical in parentheses.

The following are a few important radicals and their valences. They should be memorized, otherwise you will not be able to write the formulas of many familiar compounds.

Radical	*Formula*	*Valence*
Sulfate	SO_4	2
Carbonate	CO_3	2
Bicarbonate	HCO_3	1
Nitrate	NO_3	1
Chlorate	ClO_3	1
Hydroxide	OH	1
Ammonium	NH_4	1

The following formulas can now be easily derived from the information given in the valence tables:

Sodium carbonate	Na_2CO_3
Calcium bicarbonate	$Ca(HCO_3)_2$
Copper nitrate	$Cu(NO_3)_2$
Potassium chlorate	$KClO_3$
Calcium hydroxide	$Ca(OH)_2$
Ammonium sulfate	$(NH_4)_2SO_4$

What Are Equations? Equations express chemical reactions in their most concise form. Both starting substances (or ***reactants***) and final substances (or ***products***) are written as formulas. We must bear in mind that no atoms are destroyed in a chemical reaction and therefore there must be the same total number of atoms in the products as in the reactants. That is why reactants and products can be equated and why the concise statement of a reaction is called an equation. Let us now write equations for some of the reactions already considered:

(1) *Burning of carbon*

carbon + oxygen → carbon dioxide

Notice that reactants and products are usually joined by an arrow (→) which means "yields." Bearing in mind that an oxygen molecule is O_2, and a carbon dioxide molecule is CO_2, the equation for the reaction is

$$C + O_2 \rightarrow CO_2$$

(2) *Burning of magnesium*

magnesium + oxygen → magnesium oxide

A symbolized statement for this reaction is

$$Mg + O_2 \rightarrow MgO$$

This, however, is not an equation since there are two oxygen atoms on the left hand side and only one on the right hand side. The number of oxygen atoms can be balanced by taking two molecules of magnesium oxide (2MgO). This adjustment, however, upsets the balance between the magnesium atoms. Hence two atoms of magnesium (2Mg) are necessary. The balanced equation is, therefore,

$$2Mg + O_2 \rightarrow 2MgO$$

Notice that there are two atoms of magnesium and also two atoms of oxygen in both the reactants and products.

(3) *Burning of hydrogen*

hydrogen + oxygen → water vapor

A symbolized statement is

$$H_2 + O_2 \rightarrow H_2O$$

To balance the equation we need 2 molecules of H_2O and 2 molecules of H_2. Or,

$$2H_2 + O_2 \rightarrow 2H_2O$$

(4) *Preparation of oxygen*

potassium chlorate → potassium chloride + oxygen

Using formulas for these substances we have

$$KClO_3 \rightarrow KCl + O_2$$

There are more oxygen atoms in the reactant than in the product. The balanced equation is

$$2KClO_3 \rightarrow 2KCl + 3O_2$$

This is not an easy equation to balance. It is worthwhile to spend a little time on this equation to make sure that you understand the intermediate steps in the balancing process.

(5) *Preparation of hydrogen*

zinc + sulfuric acid →
zinc sulfate + hydrogen

$$Zn + H_2SO_4 \rightarrow ZnSO_4 + H_2$$

The symbolized equation shows clearly that the metal zinc displaces the hydrogen in sulfuric acid. Moreover, it shows that, in doing so, a salt is formed. Indeed, this reaction enables us to define that important group of substances called salts.

Salts. ***A salt is the substance formed when the hydrogen of an acid is replaced by a metal.*** If the hydrogen of sulfuric acid is replaced, a sulfate is formed. Similarly, a ***nitrate*** is a salt of nitric acid, a ***chloride*** is a salt of hydrochloric acid, and a ***carbonate*** is a salt of carbonic acid. The acid-salt relationship is shown in the following table.

Acid	*Formula*	*Corresponding Salt*	*Example*
Sulfuric	H_2SO_4	Sulfate	Na_2SO_4
Nitric	HNO_3	Nitrate	$Cu(NO_3)_2$
Hydrochloric	HCl	Chloride	$MgCl_2$
Carbonic	H_2CO_3	Carbonate	$CaCO_3$

We have now discussed acids, salts, formulas, and equations. A good way to tie all these terms together, and to illustrate their importance in elementary chemistry, is to prepare carbon dioxide and to examine some of its properties.

Carbon Dioxide. Can you think of a convenient way to prepare carbon dioxide? You may recall that in the acid test for marble, limestone or calcite (see page 47) carbon dioxide was evolved. Let us repeat this experiment by placing a few chips of marble in a test tube and pouring a little dilute hydrochloric acid into the tube. There is a brisk effervescence and a gas comes off. To write an equation for the reaction we must remember that, whenever hydrochloric acid reacts, a chloride is formed. In this case it must be calcium chloride ($CaCl_2$). Therefore, the equation might be

$$CaCO_3 + 2HCl \rightarrow CaCl_2 + H_2CO_3$$

This equation shows that carbonic acid (H_2CO_3) is formed. This is probably what happens but it is only a first step. Carbonic acid is an unstable compound and most of it decomposes as soon as it is formed. This second reaction accounts for the formation of carbon dioxide. It is

$$H_2CO_3 \rightarrow H_2O + CO_2$$

However, in writing the equation for the complete reaction we usually mention only the reactants and the ***final*** products. Or,

$$CaCO_3 + 2HCl \rightarrow CaCl_2 + H_2O + CO_2$$

Let us now collect a few bottles of the gas and discover some of its properties.

Demonstration

18-1. Preparation of carbon dioxide

Cover the bottom of a bottle with pieces of marble and then arrange the apparatus as shown in Fig. 18-1. Pour some dilute hydrochloric acid down the thistle tube. A vigorous reaction takes place. Notice that carbon dioxide is collected by displacing air, not water. Hold a burning taper near the mouth of the bottle. If the taper is extinguished, the bottle is full of carbon dioxide.

Now remove the bottle and pass the carbon dioxide into a test tube half full of water in which a few pieces of blue litmus paper have been placed. Notice that the litmus paper slowly turns red.

Now remove the test tube of water and let the gas bubble through lime water in another test tube. Notice that the lime water turns milky.

Properties of Carbon Dioxide. The above experiments illustrate important properties of carbon dioxide that we shall refer to frequently in later chapters.

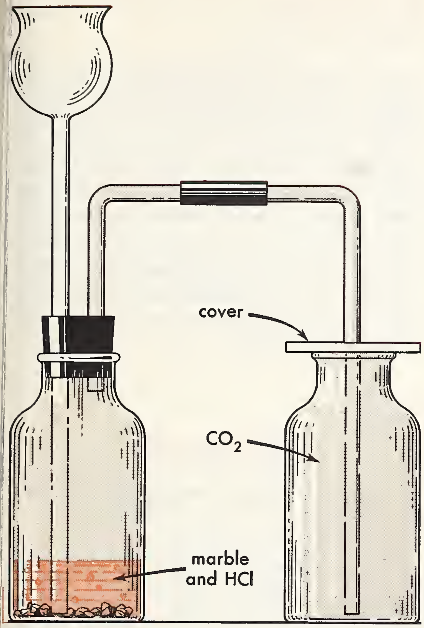


Fig. 18-1. Carbon dioxide is prepared by pouring an acid on to marble.

1. *Carbon dioxide is heavier than air.*

Because carbon dioxide is heavier than air it can be collected by displacing air. That is to say the heavy carbon dioxide is led to the bottom of the bottle and the lighter air is forced up and out of the top of the bottle.

2. *Carbon dioxide reacts with water to form an acid.*

A substance that turns blue litmus red is an acid, carbonic acid in this case. The reaction is

$$H_2O + CO_2 \rightarrow H_2CO_3$$

Carbonic acid is a weak acid. It is not nearly as active chemically as strong acids like sulfuric, hydrochloric, and nitric. The blue litmus test is one of the ways to tell whether an acid is weak or strong. A weak acid changes blue litmus to red slowly; a strong acid changes it instantly.

3. *Carbon dioxide turns lime water milky.*

If lime (CaO) is shaken in water it reacts with water to form calcium hydroxide,

$$CaO + H_2O \rightarrow Ca(OH)_2$$

Calcium hydroxide is not very soluble and only a small proportion of it dissolves in water. This dilute solution is called lime water.

Carbon dioxide reacts chemically with lime water to form calcium carbonate,

$$Ca(OH)_2 + CO_2 \rightarrow CaCO_3 + H_2O$$

Calcium carbonate, you will recall, is almost insoluble. That is why it appears as a fine milky precipitate. The exhaled breath contains about 4 per cent of carbon dioxide. For this reason, a milky precipitate is formed by breathing into lime water.

Erosion of Rocks. In Chapter 7 we discussed the erosion of limestone rocks. The question that now arises is, If limestone ($CaCO_3$) is almost insoluble, why is it eroded by running water? The dissolving of limestone rock is made possible because running water is usually a very dilute solution of carbonic acid. You will recall that vegetation releases carbon dioxide when it decomposes. Some of the carbon dioxide reacts with water to form carbonic acid and the acid, in turn, reacts with limestone to form calcium bicarbonate,

$$CaCO_3 + H_2CO_3 \rightarrow Ca(HCO_3)_2$$

Calcium bicarbonate is much more soluble than calcium carbonate. That is why limestone rocks are slowly weathered and why large subterranean caverns are formed in limestone regions. Calcium bicarbonate also makes water hard; it reacts with soap to form a scum. If hard water is boiled calcium bicarbonate is decomposed as shown in the following equation:

$$Ca(HCO_3)_2 \rightarrow CaCO_3 + H_2O + CO_2$$

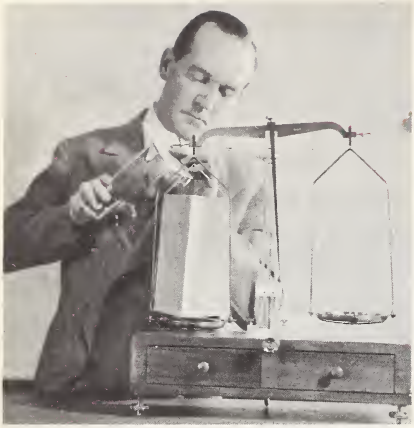


Fig. 18-2. As carbon dioxide is poured into the paper bag, the balance registered an increase in weight.

In this way the insoluble calcium carbonate is formed. Indeed, this is the objectionable rocky substance deposited as a scale in boiler tubes.

Calcium bicarbonate is, of course, carried in solution to the sea. This is the compound that sea animals use in building up their hard shells. That is to say, these animals are able to extract the insoluble calcium carbonate from a solution of the bicarbonate.

Carbon Dioxide as a Fire-extinguisher. Carbon dioxide has two properties that make it useful as a fire extinguisher. First it is about one and a half times as dense as air. Its heaviness is an advantage in extinguishing a fire because the gas will displace air and smother a flame. The following experiment shows that carbon dioxide is denser than air and flows like water.

Demonstration

18-2. Carbon dioxide is denser than air

Balance a paper bag on a chemical balance or a triple beam balance. Pour carbon dioxide into the bag as shown in Fig. 18-2. The balance registers an increase in weight as the air is displaced.

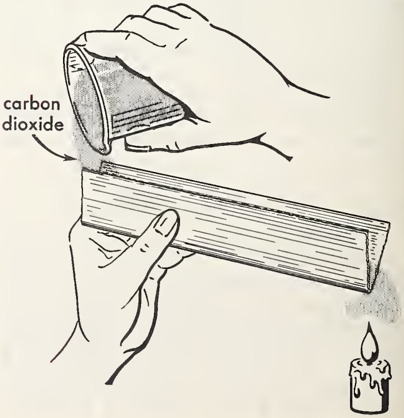
The other useful property of carbon dioxide is its chemical inactivity. The structural formula for carbon dioxide is $O{=}C{=}O$. This shows that it cannot combine with oxygen because the carbon atom already has as much oxygen as it can hold; or, expressed another way, carbon dioxide cannot burn because it is a product of combustion. Its effectiveness as a fire extinguisher is shown in the following experiment.

Demonstration

18-3. Carbon dioxide extinguishes candle flames

Fill a large beaker with carbon dioxide and cover it with a piece of cardboard. Hold the groove of a paper hand towel so that its lower end is near a candle flame. Displace the cardboard a little and pour the carbon dioxide into the upper part of the groove. The candle flame is immediately extinguished as the heavy gas covers it. (See Fig. 18-3.)

Fig. 18-3. Carbon dioxide poured down the trough extinguishes a candle flame.



Things to Remember

The ***valence*** of an element is the number of bonds by which one atom of this element can attach itself to other atoms.

A ***radical*** is a group of two or more elements that usually behaves like a single element during a chemical reaction.

Of the following ***endings, -um*** usually indicates a metal, ***-ide*** means there are two elements in the compound, ***-ate*** means that there are three elements (including oxygen) in the compound.

In an equation the number of atoms in the reactants equals the number of atoms in the products.

A ***salt*** is the substance formed when the hydrogen of an acid is replaced by a metal.

Questions

GROUP A

1. State the meanings of the following endings: (*a*) -ide, (*b*) -ate.
2. Write symbols for (*a*) a hydrogen atom, (*b*) an oxygen molecule.
3. State the meaning of the terms (*a*) reactant, (*b*) product.
4. (*a*) What is a salt? (*b*) Name the salts formed from the following acids: (i) nitric, (ii) carbonic, (iii) sulfuric.
5. (*a*) Name a strong acid and a weak acid. (*b*) What is the effect of these acids on litmus paper?
6. Name two properties of carbon dioxide that make it useful as a fire extinguisher.

GROUP B

7. Write formulas for: (*a*) hydrogen sulfide, (*b*) copper chloride, (*c*) sodium oxide, (*d*) aluminum oxide, (*e*) carbon chloride, (*f*) potassium sulfide, (*g*) phosphorus bromide.
8. Write formulas for: (*a*) sodium bicarbonate, (*b*) iron nitrate, (*c*) magnesium sulfate, (*d*) calcium chlorate, (*e*) zinc carbonate, (*f*) ammonium carbonate, (*g*) copper hydroxide.
9. Write balanced equations for the following reactions: (*a*) hydrogen + oxygen → water vapor, (*b*) zinc + sulfur → zinc sulfide, (*c*) phosphorus + oxygen → phosphoric oxide, (*d*) aluminum + sulfuric acid → aluminum sulfate + hydrogen.
10. Write balanced equations for: (*a*) The preparation of oxygen from potassium chlorate, (*b*) The reaction between water and carbon dioxide, (*c*) The reaction between carbon dioxide and lime water.
11. Explain how a water solution of carbon dioxide causes the erosion of limestone rocks.
12. (*a*) Draw a labeled diagram of the apparatus used for the preparation of carbon dioxide. (*b*) Write a word equation and a balanced symbolized equation for the reaction.

Things to Do

To show carbon dioxide is heavier than air · Cut a thin, light rod of wood about three feet long. Bore a hole through the center of the rod, push a metal pin through the hole, and support the rod by the pin on a wooden frame.

Make 2 similar rectangular bags of stout paper (about 6″ x 6″ x 9″). Attach a piece of string from corner to corner of each of the bags. Suspend the paper bags by the strings near the ends of the rod and adjust them back and forth until the rod is horizontal.

Prepare two or three bottles of carbon dioxide. Pour the carbon dioxide from each bottle into one of the paper bags. The rod acts like a see-saw. The lever arm supporting the carbon dioxide bag falls; the bag of air on the other arm rises.

19

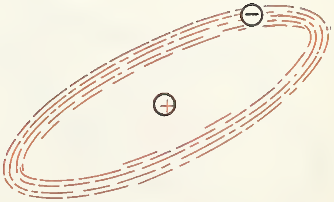
THE STRUCTURE OF ATOMS

All matter is composed of atoms, incredibly large numbers of them. Until relatively recent times it was believed that atoms were tiny, solid, indivisible particles, an idea that was bequeathed to us by the Greek philosophers some 2000 years ago. Indeed, the word atom is derived from the Greek ***atomos,*** a word which means ***indivisible.*** By the turn of the present century scientists were sure that this concept was wrong. There was evidence to show that all atoms consist of two parts, a positively charged core (or nucleus) and a shell composed of negative electrons. In 1910, Lord Rutherford, a British scientist, compared an atom to the planetary system in which the positive nucleus corresponds to the sun and the whirling, high-speed electrons to the planets.

What Are Electrons? ***Every electron is a particle of negative electricity.*** The number of electrons in the shell is the same as the number of positive charges in the nucleus. The electronic charge therefore neutralizes the positive charge on the nucleus so that an atom, under normal conditions, is neutral. The lightest element is hydrogen. A hydrogen atom has a nucleus consisting of a single positive charge and therefore there is only one revolving electron. (See Fig. 19-1.) The heaviest natural element is uranium. Uranium atoms have 92 positive charges in a nucleus which is surrounded by a "cloud" of 92 whirling electrons.

Scientists have determined the weight of electrons. They are exceedingly light—the weight of an electron is only about 1/1800 as heavy as a hydrogen atom. Nevertheless, these tiny negatively charged particles can do many things. If electrons flow through a wire they constitute an electric current—about 6 billion billion (6,000,000,000,000,000,000) of them pass through the filament of a 100-watt light bulb every second. If they sweep across a cathode-ray tube they may form a television picture. And, finally, they take part in all chemical reactions. Indeed chemical properties of atoms are

Fig. 19-1. A hydrogen atom consists of a single proton and one revolving electron.



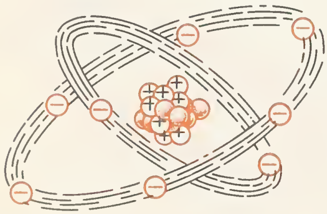
determined by the number and distribution of revolving electrons. If there are 29 electrons whirling round a nucleus the atom is copper; if there are 26 electrons, the atom is iron; if there are 82 electrons the atom is lead.

What Is a Nucleus? Scientists have even explored the nuclei of atoms and have discovered that they consist of two kinds of particles, protons and neutrons. Protons and neutrons have about the same weight (both are about 1800 times heavier than an electron), but electrically their properties are quite different. ***A proton carries a positive charge of electricity.*** The nucleus of a hydrogen atom consists of a single proton; the nucleus of a uranium atom contains 92 protons.

A neutron, on the other hand, is a particle without charge—a ***neutral*** particle as the name suggests. ***An atomic nucleus is therefore an aggregate of a certain number of protons and neutrons.*** It follows that ***the total number of particles in a nucleus constitutes the atomic weight of that particular element,*** the weight of the electrons being negligibly small. The nucleus of an oxygen atom, for instance, consists of 8 protons and 8 neutrons (see Fig. 19-2) and hence its atomic weight is 16. On the other hand, the nucleus of a uranium atom does not consist of equal numbers of protons and neutrons. Instead, it is an aggregate of 92 protons and 146 neutrons, so that its atomic weight is 238.

It should now be apparent that almost all the mass of an atom is concentrated in the nucleus. Yet Rutherford discovered that a nucleus is about 10,000 times smaller in size than the atom itself. Or, expressed another way, if an atom could be magnified to the size of a house, the nucleus would be no larger than a grain of sand. With this picture in mind it is clear that an atom is mostly empty space. This important concept is illustrated in the picture of an oxygen atom, Fig. 19-3. Here the tiny nucleus has been enlarged so that its diameter measures 500 yards. Notice that the paths of the outer electrons in this magnified atom extend from Halifax to Vancouver, a distance of about 3000 miles.

Fig. 19-2. An oxygen atom has 8 protons and 8 neutrons in the nucleus, and 8 revolving electrons.



The Periodic Chart Tells the Atomic Structure. Notice the periodic chart placed inside the back cover of this book. You will see that there are 102 different elements arranged in numerical order. For each element the periodic chart gives the following information: above the atomic symbol is the atomic number, below the symbol is the atomic weight, and in a column to the left of the atomic symbol is shown the arrangement of the electrons.

The atomic number tells us the number of electrons, all outside the nucleus. It also tells us the number of protons in the nucleus, since the atom is neutral, and it has an equal number of positive charges (on the protons) and negative charges (electrons). The atomic weight tells us the total number of protons plus neutrons, since the weight of the electrons is very small, and the mass of the atom is thus made up almost completely by the mass of the protons and neutrons in the nucleus.

Now with this information we are ready to draw a diagram which will give us some idea of the structure of the atom. We must remember, however, that the picture we draw is not what the atom would look like if we could see it, but is just a convenient way of illustrating a number of important properties of the atom. The correct picture of the atom is not known: as one learns more about atoms, one finds

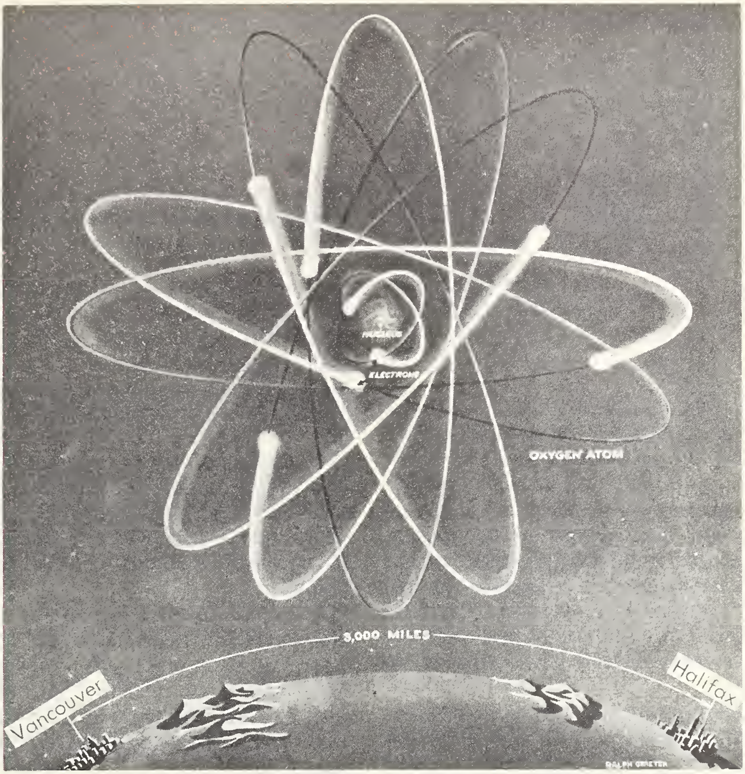
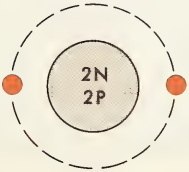


Fig. 19-3. An atom is mostly empty space. If the width of the nucleus of an oxygen atom could be enlarged to 500 yards, the diameter of the orbits of the outer electrons would be about 3000 miles.

that it is hard to imagine what this picture could look like, and instead one writes mathematical equations which describe the motion of the electrons.

A Hydrogen Atom. A hydrogen atom is the simplest of all atoms. It has an atomic weight of 1.008, and atomic number, 1. The atomic number tells us that there is one proton in the nucleus, and, since all atoms are electrically neutral, there is a single electron which revolves about the nucleus, as shown in Fig. 19-1. Since the atomic weight is approximately 1, and since we know that there must be one proton in the nucleus, we can conclude that there are no neutrons in the nucleus. A hydrogen nucleus is in fact the only nucleus that contains no neutrons.

Fig. 19-4. A helium atom has 2 protons and 2 neutrons in the nucleus, and 2 revolving electrons.



A Helium Atom. Although helium is next to hydrogen in lightness, helium atoms are considerably more complex than hydrogen atoms. The atomic number of helium is 2, and its atomic weight is 4.003. The atomic number tells us that there are

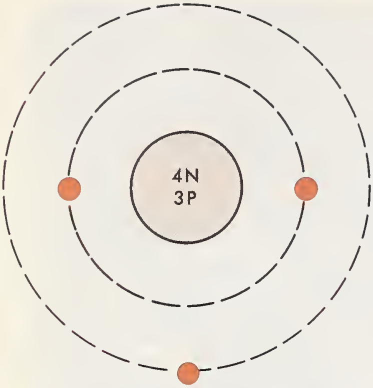


Fig. 19-5. The electrons of a lithium atom are arranged in two shells, 2 in the inner shell and 1 in the outer.

two electrons in the shell of a helium atom, and two protons in its nucleus. However, we know that its atomic weight is approximately 4. To account for this atomic weight, then, we must conclude that there are two protons and two neutrons in the nucleus. This is shown in Fig. 19-4.

Helium atoms are chemically inactive; they form no compounds. This 2-electron arrangement is therefore a stable structure. In other words, the next element, lithium, must arrange its extra electron in a second, or outer, orbit or shell.

A Lithium Atom. The atomic number of lithium is 3 and its atomic weight is 6.940. The lithium atom must then have three external electrons. To account for the atomic weight of approximately 7, the nucleus must contain 3 protons and 4 neutrons. Notice in Fig. 19-5 that the electrons are arranged in two shells, two electrons in the inner shell and one in the outer shell. Again notice, in the periodic chart, that this is indicated in the column to the left of the atomic symbol. The column shows first the number 2, which tells us there are two electrons in the inner shell, and then the number 1, which tells us the number of electrons in the outer shell. The lone electron in the outer shell is very loosely held, and takes part in all chemical reactions of the element lithium.

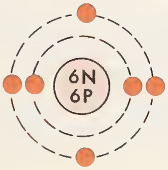


Fig. 19-6. The atomic weight of a carbon atom is 12. Why?

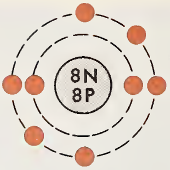


Fig. 19-7. The 8 electrons of an oxygen atom are arranged in two shells, 2 in the inner shell and 6 in the outer one.

A Carbon Atom. A carbon atom has 6 protons and 6 neutrons in its nucleus. Therefore there must be 6 revolving electrons outside the nucleus, 2 in the inner shell and 4 in the outer as shown in Fig. 19-6. The atomic weight of carbon is, of course, 12.

An Oxygen Atom. The nucleus of an oxygen atom contains 8 protons and 8

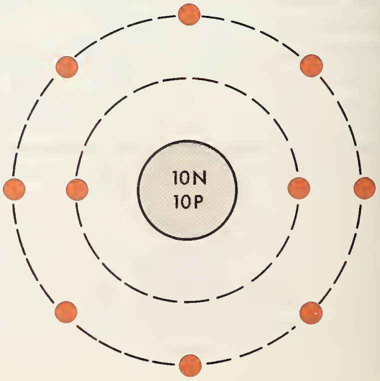


Fig. 19-8. A neon atom has the stable structure of 8 electrons in its outer shell.

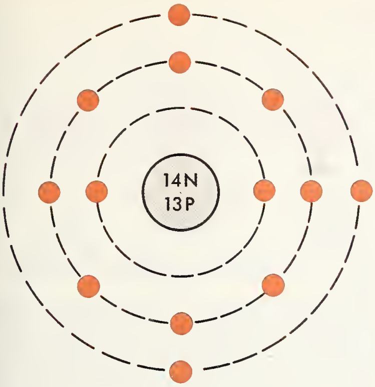


Fig. 19-9. An aluminum atom has 13 electrons: 2 in the first shell, 8 in the second, and 3 in the third.

neutrons as shown in Fig. 19-7. What is its atomic weight? Notice that the 8 electrons are arranged in two shells, 2 in the inner one and 6 in the outer one. Now refer again to the picture of the oxygen atom on page 182 and notice particularly the orbits of the inner and outer electrons.

A Neon Atom. Neon is a gas. We know that neon atoms in a sealed tube give a red glow if a high voltage is applied at the ends of the tube, an effect made use of in advertising signs. The nucleus of a neon atom consists of 10 protons and 10 neutrons. Therefore there are 10 electrons revolving at high speed about each nucleus. How are these electrons arranged? Again the answer is 2 in the inner shell and 8 in the outer one. We have already stated that the outermost electrons govern the chemical properties of an atom. But neon has no marked chemical properties. That is to say, it is an inert element; it takes no part in any chemical reactions. Its chemical stability is due to the fact that there are 8 electrons in its outer shell. This is a stable structure. Indeed no atom can have more than 8 electrons in its outermost shell. Hence, the next heavier atom than neon starts a third shell for its outermost electrons.

An Aluminum Atom. The nucleus of an aluminum atom contains 13 protons and 14 neutrons. Hence there must be 13 revolving electrons. How are they arranged? Two in the first shell, 8 in the second, and 3 in the third, an arrangement shown in Fig. 19-9.

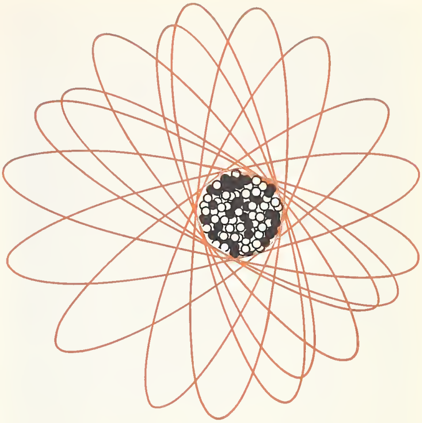


Fig. 19-10. A uranium atom is the heaviest natural element. Its nucleus contains 92 protons and 146 neutrons.

Copper and Lead Atoms. As the atomic weight of an atom increases, its nucleus becomes more massive and the paths of the revolving electrons more complex. The nucleus of a copper atom contains 29 protons and that of a lead atom 82 protons. It therefore follows that there are 29 electrons revolving around the copper nucleus and 82 around the lead nucleus. In addition to the protons there are 34 neutrons in the copper nucleus and 126 neutrons in the lead nucleus. What are the atomic weights of copper and lead?

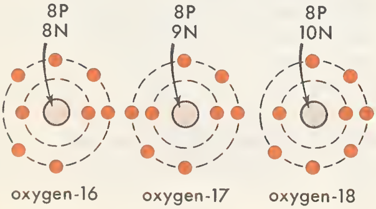
A Uranium Atom. The heaviest naturally occurring atom is that of uranium with an atomic weight of 238. The massive nucleus of this element contains 92 protons and 146 neutrons. Indeed the nucleus is too massive to remain stable and it decomposes spontaneously, a radioactive process discussed later in the chapter.

ISOTOPES

At one time it was thought that all atoms of the same element were exactly alike. It was believed, for example, that all oxygen atoms have a weight of 16. We now know, however, that the atomic weight of most oxygen atoms is 16, but that for some the atomic weight is 17, and for a few it is 18. How can this be? All oxygen atoms have the same chemical properties and therefore they all have the same number (8) of revolving electrons. This, in turn, means that all have the same number of protons (8) in the nucleus. Hence the only way to account for the differences in weight is by differences in the number of neutrons. A difference in the number of neutrons does not affect the number of electrons so that the chemical properties of the atoms are in no wise changed. The nucleus of oxygen-16 contains 8 neutrons; oxygen-17 contains 9 neutrons; oxygen-18 contains 10 neutrons. These different atomic forms of oxygen are called ***isotopes.***

Isotopes are simply atoms of the same element but with different weights. Almost every element has two or more natural isotopes: carbon has two, nitrogen two, oxygen three, and even hydrogen has three. An isotope of hydrogen called deuterium or heavy hydrogen has a double-weight nucleus consisting of a proton and a neutron as shown in Fig. 19-12. Tritium is the third isotope of hydrogen, its nucleus consisting of a proton and 2 neutrons.

Fig. 19-11. The isotopes of oxygen are oxygen-16, oxygen-17, and oxygen-18.



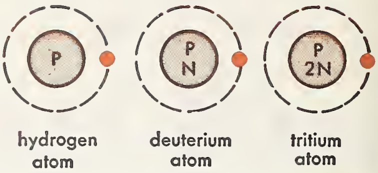


Fig. 19-12. The three isotopes of hydrogen.

Or again there are two isotopes of copper. About two-thirds of copper atoms have an atomic weight of 63 and the rest have an atomic weight of 65. Can you now draw diagrams of the nuclei of the isotopes of copper?

Stability of Isotopes. Since most elements have only a few stable isotopes, it follows that there are only a few stable combinations of protons and neutrons. We shall read later, in Chapter 33, that it is a fairly easy matter to add an extra neutron to the nucleus of an atom. But when a nucleus is penetrated by a neutron it usually becomes unstable and disintegrates.

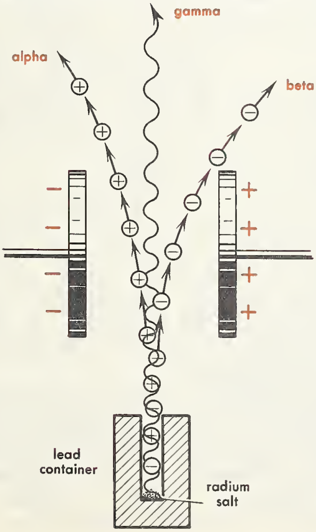
Probably the most interesting isotopes are those of uranium (atomic weights 235 and 238), usually called U-235 and U-238. U-238 is far the more plentiful isotope. Actually natural uranium is a mixture of 99.3 per cent U-238 and 0.7 per cent U-235. U-235 is the explosive kind. That is to say, its nucleus splits into two nearly equal parts if it is penetrated by a single neutron, a process called ***fission.*** U-238, on the other hand, does not explode in this way; its nucleus can absorb a neutron without splitting. This marked difference between the two isotopes made it necessary to separate them before the U-235 atomic bomb could be made.

Radioactivity. It was the study of radioactivity that provided most of the clues in the early investigation of atomic structure. Radioactivity was discovered by accident in 1896, when a French scientist named Becquerel placed a lump of uranium ore on a box of photographic plates.

When the plates were later developed, they were all found to be fogged. Why? This to Becquerel was an intriguing question. He concluded that uranium atoms shot off invisible radiations capable of penetrating the cardboard box and fogging the photographic plates. These radiations are a form of energy. Where did the energy come from? Was it supplied by atoms of uranium? Becquerel strongly heated the ore and then cooled it to a low temperature. But the radiations escaped at the same rate and there was nothing he could do to change this rate. Becquerel had stumbled upon a new and fascinating field for scientific research.

Although radioactivity was discovered by Becquerel, the metal uranium had been known for more than a century. Actually it was discovered by a German chemist named Klaproth in 1789. The chief use of uranium compounds had been to add color (ruby, orange or yellow) to glass and porcelain. The stained glass windows of many European cathedrals, for example, are colored by uranium compounds. And the uranium used for this purpose was extracted from pitchblende, a hard, black uranium ore found in Czechoslovakia.

Fig. 19-13. Alpha and beta particles are deflected by an electric field, but not gamma rays.



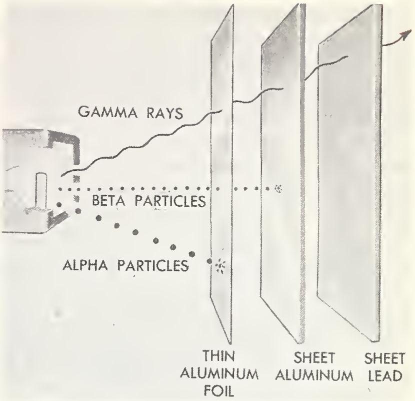


Fig. 19-14. Alpha particles are stopped by thin aluminum foil; beta particles are stopped by aluminum sheet one centimeter thick; gamma rays will penetrate sheet lead five centimeters thick.

The Curies. Marie Curie and her scientist husband, Pierre, close friends of Becquerel, began work in this new field. First the Curies discovered the remarkable fact that pitchblende was more radioactive than uranium itself. How could that be? Apparently the ore contained an element more radioactive than uranium and the Curies set about to find it. This tedious and strenuous task kept them occupied for nearly four years. Finally success came. From one ton of pitchblende they obtained a minute quantity (about the size of a pinhead) of a new element which was about 1000 times as radioactive as uranium. This new element they called ***radium.***

Radiations from Radium. What kind of radiations are emitted from radium? From what part of the atom do they come? These questions were investigated by the British scientist, Lord Rutherford. He found three different kinds of radiations. Some were positively charged, some were negatively charged, and some were not charged at all. He called them alpha, beta, and gamma rays.

Alpha rays are really positively charged particles. These particles are composed of two protons and two neutrons which break away from the larger nucleus; they are, in fact, the nuclei of helium atoms.

Beta rays are negatively charged particles; they are electrons. But how can electrons escape from a nucleus that consists only of protons and neutrons? A neutron can be changed to a proton and an electron. Indeed, this is one of the changes in the radioactive process. When the change takes place the released electron is shot out of the nucleus at high speed, almost at the speed of light.

Gamma rays, on the other hand, are not particles; they do not consist of protons, neutrons or electrons. They are electromagnetic radiations ***similar to X rays*** and travel with the speed of light. They are highly penetrating and can pass through a steel block a foot thick. They can also penetrate human tissue, so that prolonged exposure to them is highly dangerous.

Radioactive Disintegration. Every known ore of uranium contains, besides uranium, such radioactive elements as thorium, radium, radon, polonium and bismuth, as well as lead that is not radioactive. Why this mixture of elements? Let us remember that both alpha and beta particles come from the nuclei of atoms. The "atomic" weight of an alpha particle is 4, since it consists of two protons and two neutrons. Therefore, if a nucleus shoots off an alpha particle an entirely new nucleus is formed; that is, a new element results from the loss of an alpha particle. Similarly, before a nucleus can eject a beta particle a neutron must be changed to a proton and an electron. Thus when this electron is emitted, the number of protons in the nucleus is increased by one and again a new element is thereby formed.

It is now apparent that the nuclei of radioactive atoms are unstable. Starting with the heaviest element, uranium, 14 nuclear disruptions occur before a stable nucleus is obtained. The stable nucleus is that of lead, and here radioactivity stops. When the Curies set out to find the potent radioactive element in uranium ore they little realized that the ore contained 14 different kinds of atomic nuclei. Even today the only way that radium is obtained is by the laborious extraction of it from a uranium ore containing only a minute fraction of radium. That is why radium is so expensive—it costs about $25,000 per gram—and why it is so fantastically scarce—the total weight of all the radium in the world does not amount to more than 5 pounds.

The Age of Radioactive Elements. We have said that when a radioactive atom blows up it releases some of its energy and forms an entirely different element of less mass. The "lives" of radioactive elements vary enormously—some are exceedingly long, others are exceedingly short. Scientists speak of the "half-lives" rather than the "lives" of radioactive elements. ***The half-life of a radioactive element is the time required for half of its atoms to disintegrate.***

Uranium has the longest half-life among the radioactive elements, approximately 4½ billion years. It is generally agreed that the age of the earth is at least three billion years. It therefore follows that the amount of uranium now in the earth is about two-thirds the amount present when the earth was formed. In other words, only about one-third of the original amount of uranium has disintegrated into other elements.

The half-life of radium is about 1600 years. That is to say, one-half of a lump of radium (say, one gram of it) disintegrates in 1600 years; one-half the remainder (½ gram) disintegrates in 1600 years; one-half the remainder (¼ gram) disintegrates in the next 1600 years; and so on indefinitely.

Uses of Radium. Radium is used in the treatment of cancer. Its radiations penetrate human tissue and attack malignant growths. However, radon, a gaseous decay product of radium, is usually preferred for this work. Radon gas, the

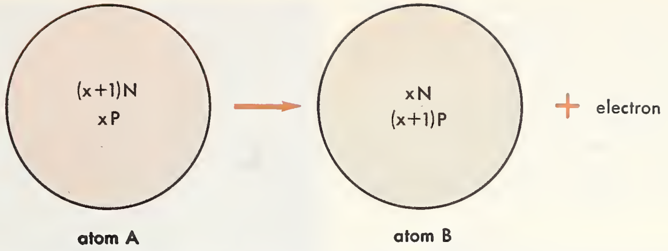


Fig. 19-15. Radioactive disintegration. A neutron in atom A changes to a proton and an electron. The electron (beta particle) is ejected thereby forming a new atom B.

heaviest of all gases—seven times denser than air—is pumped into thin-walled glass tubes which are sealed and applied to the diseased tissue. Radon has a very short half-life of only about four days and, for this reason, is more radioactive than radium. However, some radioisotopes made in a reactor are more suitable for therapeutic purposes than radium and radon in the treatment of cancer.

Radium compounds are also used in making luminous paint. A tiny amount of a radium compound (only a few cents worth) is added to paint which contains a fluorescent substance such as zinc sulfide. Radiations from the radium cause the zinc sulfide to fluoresce so that if the paint is applied to the dials of watches, clocks, compasses, or airplane instruments they can be seen in the dark.

Some Geologic Applications of Radiations. Radioactive substances in the earth's crust act as time clocks and give reliable information concerning the age of the earth. A uranium ore always contains some lead, which is a decay product of the uranium. The amount of lead in a piece of ore increases steadily with the length of time from the formation of the ore to the present day. The rate at which uranium disintegrates is known, and therefore the age of the rock can be computed from the relative weights of lead and uranium in the ore. Some rocks examined in this way have been found to be at least two and a half billion years old.

Then again every radioactive rock generates heat as it disintegrates. Every gram of radium, for example, emits millions of alpha particles per second. These particles are stopped by surrounding material so that their energy of motion is converted into heat. From a gram of radium heat is evolved at the rate of about 100 calories per hour, and it can be shown that the heat liberated during its complete disintegration would amount to almost two billion (2,000,000,000) calories. Some geologists believe that the earth is maintained at its present temperature by the radioactive changes in rocks, and that, without this source of heat, the earth would be too cold to support life.

Atomic Bullets from Radium. Rutherford hit upon the clever idea of using alpha particles shot off by radium to investigate the structure of atoms. By using alpha particles as "bullets" he hoped to split atoms and, in this way, to learn of their structure. The effectiveness of a bullet depends upon its mass and speed. And the fantastically high speeds of the alpha particles (great enough to encircle the earth at the Equator in less than one second) seemed to make them ideal "bullets" for bombarding atoms.

Rutherford bombarded various elements with alpha particles from a radioactive source. He noticed that when light elements such as nitrogen were bombarded, positively charged particles were invariably ejected from the bombarded nuclei. He therefore concluded that positive particles were an essential ingredient

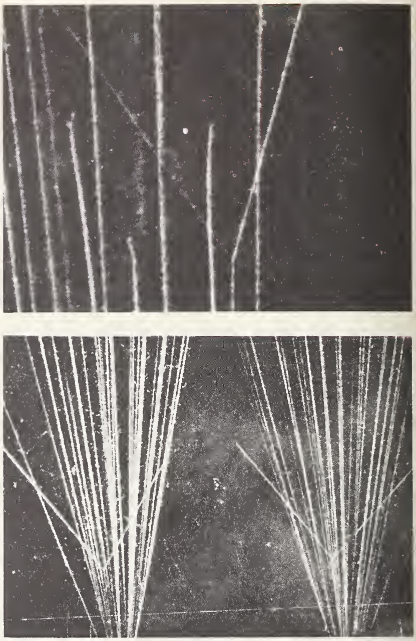


Fig. 19-16. The upper picture shows alpha particles streaming through a cloud chamber. The collision is between an alpha particle and a hydrogen nucleus. The light hydrogen nucleus is deflected to the left, but the path of the heavier alpha particle is deflected only slightly to the right.

In the lower picture an alpha particle collides with a helium nucleus. These particles have the same mass and appear to be deflected equally.

of every nucleus. He appropriately called them ***protons,*** a word derived from the Greek "protos" meaning ***first,*** the original particle in all atoms. By a suitable device (a cloud chamber), the tracks of the particles were photographed. From the millions of particles ejected by the source, only an occasional bulls'-eye was scored. When a nucleus target was hit, the alpha particle was sometimes deflected in one direction and the nucleus itself in another. Such a situation is shown in Fig. 19-16. Most alpha particles passed through the bombarded atoms without a collision, and from this observation it was concluded that atoms are mostly empty space.

The phenomena of radioactivity shed much light on the structure of atoms. And what of the energy released by radioactive atoms? Where does it come from? Scientists now know that this energy comes from matter. Radium has been converting mass into energy since the earth was formed. Is it possible to convert mass into energy on a large scale? This important question is answered in the Chapter 33.

THE PERIODIC CHART

We have already seen how the periodic chart is used to tell us the structure of the atoms. It also tells us about the properties of the elements.

You will notice that the elements are divided into groups. In each of these groups the elements have the same number of electrons in the outermost shell. Since the electrons in the outermost shells are the ones which enter into chemical reaction, then elements with the same number of electrons in the outer shells will behave in a similar manner.

In group IA lithium has one electron in the outer shell, and so has cesium. We will then expect them to react in a similar way and form compounds like LiCl and CsCl.

In group IIA we can expect calcium and radium to have similar properties, since they both have two electrons in the outer shell.

These family resemblances help us to remember the properties of the elements.

When elements react, they do so in such a way as to complete their outer shells of electrons. (For any atom other than hydrogen or helium, the outer shell can hold eight electrons.) Atoms can do this in two ways, by giving away electrons,

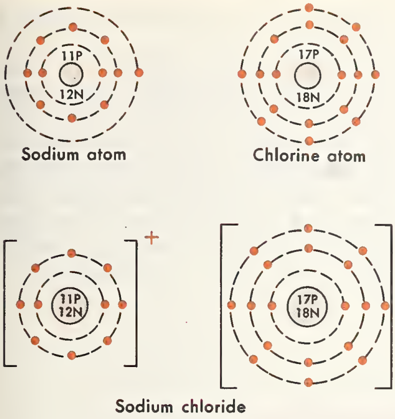


Fig. 19-17. Sodium and chlorine react to form sodium chloride.

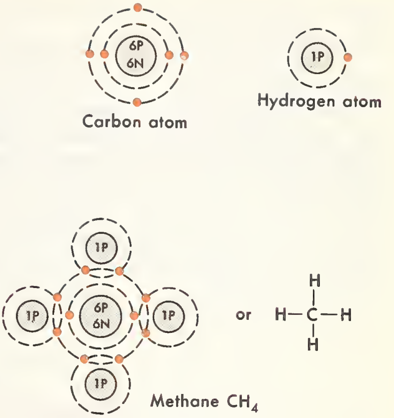
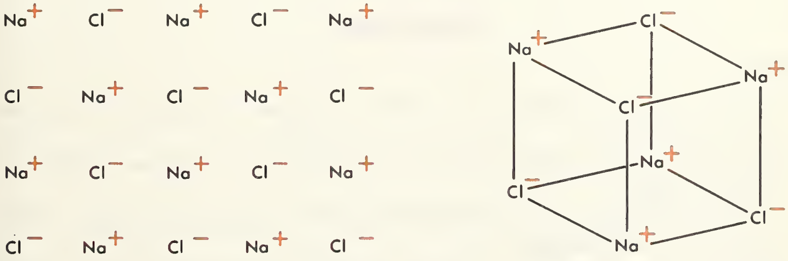


Fig. 19-18. Carbon and hydrogen react to form methane.

or by sharing electrons. Therefore there are two ways in which chemical bonds may be formed.

How Compounds are Formed by Giving Away Electrons: *Electrovalent Bonds.* When sodium reacts with chlorine it does so in such a way as to give both elements in the reaction a more suitable arrangement of electrons. The sodium atom, before it reacts, has one electron in its outer shell; the chlorine atom, before it reacts, has seven electrons in its outer shell. When they react, the sodium atom gives one electron to the chlorine atom. In this way both have eight electrons in their outer shells, which is a suitable arrangement.

All salts contain charged particles similar to the type found in sodium chloride. Notice that the Na atom has lost an electron, and so has one extra positive charge: it is now a positive ion. The chlorine has gained an extra electron. It now has one extra negative charge, and is a negative ion. The resulting sodium chloride crystal is made up of these positive and negative ions held in position by the electrostatic forces between them which



Sodium chloride in two dimensions

Sodium chloride in three dimensions, a cubic arrangement

Fig. 19-19. Sodium ions and chloride ions form a crystal structure.

form a three-dimensional lattice with an ion at each corner of a cube.

We call this type of bond an electrovalent bond.

How Compounds are Formed by Sharing Electrons: *Covalent Bonds.* Some atoms do not like to give electrons away. Instead they will only share them with other atoms. Carbon is one of these.

The compound methane is found to contain one carbon atom and four hydrogen atoms (CH_4). The structure of methane can be considered as shown in Fig. 19-18.

Notice that two electrons are shared by two atoms. The shells of the atoms overlap, and the electrons appear to be in both shells. This forms a covalent bond. Actually, although it seems elements like to have eight electrons in the outer shell (as carbon has in methane), the most important concept is that two electrons be shared to form a bond.

Things to Remember

An ***electron*** is a particle of negative electricity.

The ***nucleus*** of an atom consists of protons and neutrons.

The ***atomic weight of an element*** is the number of protons and neutrons in the nucleus of one of its atoms.

A ***proton*** carries a positive charge of electricity.

The ***weight of an electron*** is about $\frac{1}{1800}$ the weight of a proton or neutron.

Isotopes are atoms of the same element but with different atomic weights.

Natural uranium consists of 2 isotopes, U-238 (plentiful) and U-235 (rare).

U-235 is ***fissionable*** by slow neutrons.

Alpha rays are positively charged particles.

Beta rays are electrons.

Gamma rays are similar to X rays.

In ***radioactivity,*** alpha, beta, and gamma rays are emitted.

The ***half-life of a radioactive element*** is the time required for half of its atoms to disintegrate.

Questions

GROUP A

1. What is (*a*) an electron, (*b*) a proton? Compare the weights of a proton and an electron.
2. The nucleus of an atom contains 11 protons and 12 neutrons. What is its atomic weight?
3. What are isotopes of an element? Name the 3 isotopes of hydrogen. What are their atomic weights?
4. Name the two isotopes in natural uranium. In what proportion are they present?
5. What was the particular contribution of the Curies to radioactivity?
6. Name some of the uses of radium.

GROUP B

7. The atomic weight of an atom is 25, and there are 12 protons in the nucleus. How many neutrons are there in the nucleus? Draw a diagram of the atom.
8. What is meant by the term ***stable structure?*** Illustrate your answer by reference to the structures of atoms of helium and neon.
9. The atomic weights of the two isotopes of carbon are 12 and 13. Draw atomic diagrams of these isotopes.
10. What are alpha, beta, and gamma rays? Explain.
11. Explain why a new element is formed if a radioactive atom ejects a beta particle.
12. Explain why the concept of ***half-life*** is more useful in radioactivity than the concept of ***whole life.***
13. Some geologists state that without radioactivity the earth would be too cold to support life. Explain what this means.
14. What did Rutherford discover by bombarding various elements with alpha particles? Explain.

Things to Do

Alpha particles from a wrist watch · Look at the luminous dial of a wrist watch or a compass in a room that is totally dark. Examine one of the numerals closely through a powerful magnifying glass. When your eyes have become used to the darkness you will note that the numeral is not illuminated by a steady glow of light. Instead, light is emitted in flashes.

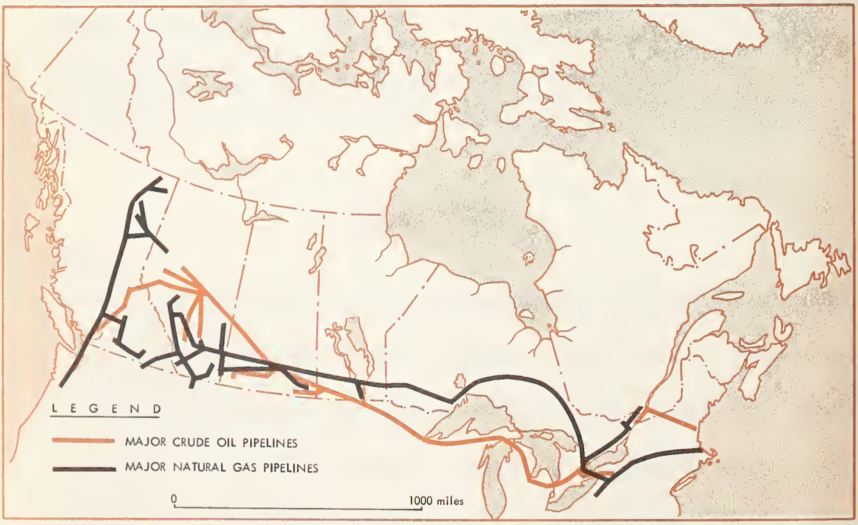
The flashes of light are scintillations emitted every time an alpha particle strikes a molecule of zinc sulfide.

20

OIL AND NATURAL GAS

The advance of technology in the world has been based upon the supply of energy available. The early technological development of Great Britain, western Europe, and the United States depended to a great extent upon their large supplies of coal. With this coal they could develop industries. Today we have many fuels available, and, in contrast to the nineteenth century when industry had to be brought to the fuel, the fuel can now be transported to the centers of industry by

Fig. 20-1. Today fuel is transported to the centres of industry. Crude oil and natural gas are Canada's major energy source.



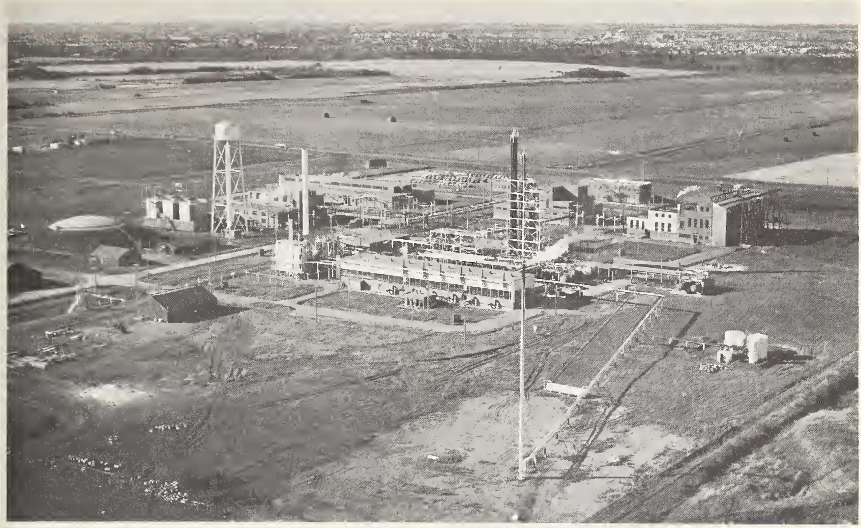


Fig. 20-2. Polythene works of Canadian Industries Limited at Edmonton. This is one of the many petrochemical industries that result from a plentiful supply of natural gas in Alberta.

land and sea carrier and by pipeline.

The important sources of energy today are coal, oil, natural gas, water power and nuclear energy. In this chapter we shall consider the fossil fuels: coal, oil and natural gas. Later you will learn something of nuclear energy.

Not only are coal, oil and natural gas sources of energy; they are also the raw materials from which many modern chemicals and plastics are manufactured. The branch of chemistry which includes the transformation of coal, oil and gas into useful compounds is called carbon or organic chemistry. Can you name one other branch of chemistry?

Coal. About one-third of all coal mined is used directly as a fuel, mainly for the generation of electrical power. The remainder is converted by destructive distillation into coke and its useful by-products. The primary products from this distillation are coke, coal gas, ammonia, coal tar, and light oil. The last two are mixtures of various chemicals, and are a valuable source of organic compounds.

Coke. Coke is nearly pure carbon. It is used to some extent for heating, but is mostly consumed in the separation of metals from their minerals. It is useful for this purpose because coke is very strong, although porous, and can support the heavy load of the minerals that are fed into a furnace, and also because the heat that coke supplies is about 13,000 to 14,000 B.t.u. per pound.

Coke may be used as a starting material for "water gas." When water gas is made, coke is burned with air until it reaches a temperature of 1000°C. Then the air is turned off, and steam is blown into the furnace. The reaction

$$C + H_2O \rightarrow CO + H_2$$

occurs. Since this reaction absorbs heat, the furnace will cool, and so, when the temperature falls, it is necessary to turn the steam off and blow more air into the coke.

Coal Gas. Coal gas is a mixture of hydrogen, 50 per cent; methane, 35 per cent; carbon monoxide, 6 per cent; and small amounts of other gases. This gas and water gas may be used for cooking and heating, but they both are also important as starting materials for organic compounds. If carbon monoxide and hydrogen are brought together in contact with a suitable catalyst under proper conditions of temperature and pressure, they can be converted into gasoline and other liquid hydrocarbons.

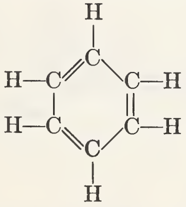
Ammonia. Before coal gas is used as a fuel, ammonia (NH_3) is removed from it. The gases are passed through a dilute solution of sulfuric acid; the product of the reaction, ammonium sulfate, forms a saturated solution,

$$2NH_3 + H_2SO_4 \rightarrow (NH_4)_2SO_4$$

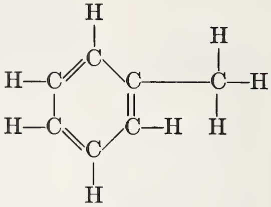
The solution is cooled and the ammonium sulfate crystallizes out. It is used mostly as a fertilizer.

Coal Tar and Light Oil. Coal tar and light oil are separated from crude coal gas, and further refined by distillation. Some of the products separated are benzene, toluene, xylene, phenol, naphthalene, and anthracene. These are basic products; they are obtained on a large scale in industry. What are they used for? What are they like chemically?

Practically all the compounds in light oil and tar are made up of hexagonal ring-shaped molecules. The basic unit of all these molecules is benzene (C_6H_6). A molecule of benzene consists of a hexagon of 6 carbon atoms to each of which a hydrogen atom is attached. Structurally it can be written as

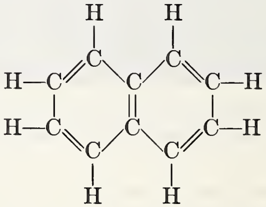


Notice that the valence of carbon is 4, and of hydrogen one. Other compounds in this series can be built up from benzene by adding carbon atoms one at a time. For example, toluene is C_7H_8 or



and xylene is C_8H_{10}.

If the number of carbon atoms increases to 10, two benzene rings fuse together to form naphthalene, $C_{10}H_8$, or



If three rings fuse together, anthracene ($C_{14}H_{10}$) is formed.

The above coal derivatives are the foundation stones of synthetic chemicals. Chemists are able to link them up in various ways or to substitute such atoms as oxygen, sulfur, nitrogen and chlorine for the hydrogen atoms. In this way a vast number of new compounds have been synthesized, some of great commercial importance.

Coal By-products in Industry. The approximate yields from the distillation of one ton of coal are

Coal gas	10,000 cubic feet (about 400 pounds)
coke	1400 pounds
tar	10 gallons (about 90 pounds)
benzene	2 gallons (about 18 pounds)

toluene	0.5 gallon (about 5 pounds)
xylene	0.25 gallon (about 2.5 pounds)
naphthalene	0.5 pounds
ammonium sulfate	24 pounds

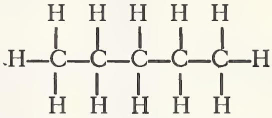
Benzene is a low-boiling aromatic liquid, the basis of numerous industrial operations. Aniline, which is manufactured from it, is a base for many dyes used in the dyeing of cloth. Then again benzene is used to make styrene, which, in turn, is used in the manufacture of synthetic rubber. Or, again, benzene may be converted into phenol which, in turn, is used to make plastics. In addition, it is used in the production of drugs, as a solvent for paint, lacquer, and varnish, and as a blend for gasoline to increase its octane rating.

A similar list of dependent industries could be made out for toluene, xylene, naphthalene, and anthracene. Indeed, the by-products of coal supply essential starting materials for the manufacture of dyes, drugs, perfumes, plastics, fibers, rubbers, explosives, fertilizers, detergents, insecticides, and even cosmetics. No wonder it has been said that the most wasteful thing to do with coal is to burn it!

Oil. Petroleum products, of course, are useful for the production of gasoline and diesel fuels to drive airplanes, automobiles, ships and locomotives, but increasing amounts of petroleum products are used in our chemical industries.

Crude oil and natural gas often occur together. If so, they are separated in a gas separator, the natural gas going to the gas treating plant, and the crude oil to the refineries.

Petroleum is made up of a large number of different hydrocarbons, that is, molecules containing carbon atoms linked together to form a chain and hydrogen atoms joined to these carbon atoms. For example, one structural formula for pentane (C_5H_{12}) is



$$\begin{array}{ccccccccccccc} & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \\ & & | & & | & & | & & | & & | & & \\ \mathrm{H} & - & \mathrm{C} & - & \mathrm{C} & - & \mathrm{C} & - & \mathrm{C} & - & \mathrm{C} & - & \mathrm{H} \\ & & | & & | & & | & & | & & | & & \\ & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \end{array}$$

Other hydrocarbons are hexane (C_6H_{14}), heptane (C_7H_{16}), etc. You will notice that the formulae for these compounds can be generalized as C_nH_{2n+2}. These are the paraffin series of hydrocarbons.

The various molecules in crude oil are separated in the refinery by distillation. The larger the hydrocarbon molecule, the higher will be its boiling point. Therefore, if we heat a mixture of hydrocarbons, the smaller or low-boiling molecules will come off first, and, if we take these molecules away at different stages of boiling, then we obtain the different fractions of crude oil. The following oil fractions are obtained:

Gasoline, a mixture of low-boiling hydrocarbons like C_5H_{12}, C_6H_{14}, C_7H_{16}, C_8H_{18}, etc., which is quite volatile;

Kerosene, a mixture of many hydrocarbons of the range $C_{12}H_{26}$, which is a good fuel for jet engines;

Fuel oil, a mixture of the range $C_{16}H_{34}$, which is used in household heating units and diesel engines;

Lubricating oil, a very high-boiling fraction which, because of this very property, can be used where an anti-fraction material which will not vaporize is required.

Asphalt, a viscous black substance that remains after the other fractions are separated.

Modifying Nature's Products. Not nearly enough gasoline is obtained by the distillation of crude oil to meet the needs of automobiles and airplanes. Indeed, unless another source of gasoline had been found, the development of the airplane would have been seriously hampered. Chemists turned to the other fractions of the distillation of crude oil to increase the gasoline supply. To do this, they

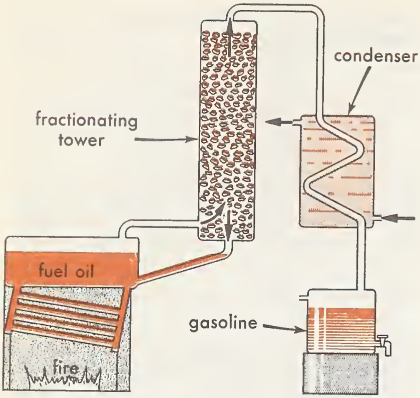
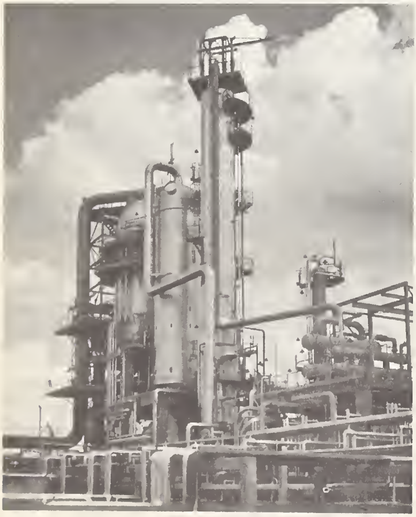


Fig. 20-3. Cracking fuel oil by high pressure and high temperature.

hoped to break down the long carbon-chain molecules of fuel oil and kerosene into shorter carbon-chain molecules of gasoline. This idea led to the industrial process of breaking carbon chains. Today, almost 50 per cent of the gasoline used is obtained by cracking.

How Are Long-Chain Molecules Cracked? Fuel oil or kerosene is subjected to a high temperature and a high pressure at the same time. The high temperature increases the speed of the molecules so that they strike each other with great force; the high pressure forces the molecules close together, thereby increasing the collisions between them. Under these conditions large numbers of long carbon chains are broken and gasoline molecules are formed.

Fig. 20-4. A catalytic cracker at Imperial Oil's Calgary refinery.



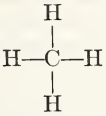
In the cracking process, vapors of fuel oil and cracked molecules pass from the boiler and boiler tubes (where the cracking takes place) to a fractionating tower as shown in Fig. 20-3. Here the unbroken molecules, which have a relatively high boiling point, are condensed and returned to the boiler for reheating. Gasoline vapor passes through the tower and is condensed further along the line. Why is gasoline vapor not condensed in the tower?

Natural Gas. Natural gas is almost always present in oil fields. Much of the natural gas, however, may be obtained from areas where there is little oil.

Natural gas obtained from gas wells is called "dry" natural gas. It may contain about 92 per cent methane, 3 per cent ethane and higher hydrocarbons, and 5 per cent nitrogen. Its heat value is about 990 B.t.u. per cubic foot. Since many of these gases have very little odor, they may be "odorized" so that when a leak occurs it can be detected before a dangerous concentration of gas accumulates. To odorize gas, compounds containing sulfur are added. These are called mercaptans. C_2H_5SH is the formula of one mercaptan.

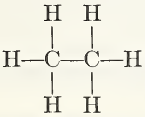
Other natural gases, called "wet" natural gas, may contain various amounts of hydrocarbons like propane, C_3H_8, and butane, C_4H_{10}. These higher-boiling hydrocarbons are separated from the methane before it is used as a fuel. They are then sold as liquid pertoleum gas (L.P. G.). The heat value for the remaining gas is then about 1160 B.t.u. per cubic foot.

Petroleum and Natural Gas in Chemical Industries. In discussing coal, we stated that benzene is the basic unit of the aromatic compounds which consist of ring-shaped molecules. The other big series of organic compounds (the aliphatics) is made up of carbon chains. Methane, CH_4, is the basic unit of the carbon chain compounds. It is the simplest of all hydrocarbons; its structural formula is



$$\begin{array}{ccccc} & & H & & \\ & & | & & \\ H & - & C & - & H \\ & & | & & \\ & & H & & \end{array}$$

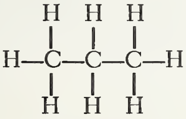
The aliphatic series can be built up from methane just as the aromatics can be built up from benzene. That is, if a carbon atom is added to each molecule, progressively longer chains are formed. For example, ethane, C_2H_6, or



$$\begin{array}{ccccccc} & & H & & H & & \\ & & | & & | & & \\ H & - & C & - & C & - & H \\ & & | & & | & & \\ & & H & & H & & \end{array}$$

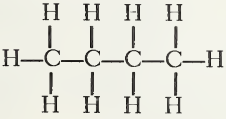
is the simplest carbon chain compound.

Propane, the next member of the series, is drawn as



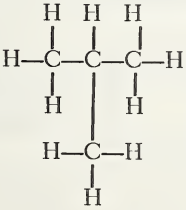
$$\begin{array}{ccccccccc} & & H & & H & & H & & \\ & & | & & | & & | & & \\ H & - & C & - & C & - & C & - & H \\ & & | & & | & & | & & \\ & & H & & H & & H & & \end{array}$$

Butane may be drawn as



$$\begin{array}{ccccccccccc} & & H & & H & & H & & H & & \\ & & | & & | & & | & & | & & \\ H & - & C & - & C & - & C & - & C & - & H \\ & & | & & | & & | & & | & & \\ & & H & & H & & H & & H & & \end{array}$$

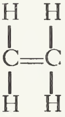
or



Compounds like these two butanes having the same formulae are called isomers. The first of these is called normal butane, the second is called isobutane. Even though isomers have the same molecular formulae, their different arrangement of carbon atoms causes them to have slightly different physical and chemical properties. For example, the boiling point of n-butane is —1°C whereas the boiling point of i-butane is —10°C. As you can see, as the number of carbon atoms increases in the series, many isomers are possible.

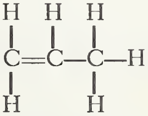
The aliphatic series includes other groups of compounds as well as the paraffin or "saturated" series. Two other groups form "unsaturated" series of hydrocarbons. In an unsaturated series, more than one bond ties two carbon atoms together. They are considered as unsaturated because the double or triple bond could bind more hydrogen atoms or other atoms to the carbon-carbon chain.

First there is the ethylene series, with a double bond, which has the group formula Cn H_2n and consists of compounds like ethylene,



$$\begin{array}{ccc} H & & H \\ | & & | \\ C & = & C \\ | & & | \\ H & & H \end{array}$$

and propylene,

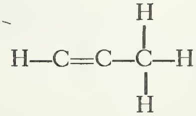


$$\begin{array}{ccccccc} H & & H & & H & & \\ | & & | & & | & & \\ C & = & C & - & C & - & H \\ | & & & & | & & \\ H & & & & H & & \end{array}$$

Then there is the acetylene series, with a triple bond, which has the group formula Cn Hn_2—$_2$ and consists of compounds like acetylene or ethyne,

$$H-C=C-H$$

and propyne,



$$\begin{array}{ccccccccc} & & & & & & H & & \\ & & & & & & | & & \\ H & - & C & = & C & - & C & - & H \\ & & & & & & | & & \\ & & & & & & H & & \end{array}$$

Write the formulae, both molecular and structural, of the unsaturated compounds containing four carbon atoms. Note the many isomers possible.

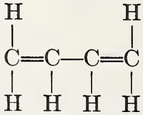
Compounds of the aliphatic series may be oxidized or burned in air to form carbon dioxide and water:

$$C_4H_8 + 6O_2 \rightarrow 4CO_2 + 4H_2O$$

Chain compounds are abundant in natural gas and petroleum and, like the coal tar derivatives, they are the starting substances for mighty industries. Chemicals derived from coal are called ***carbochemicals;*** those from petroleum are called ***petrochemicals.*** Important petrochemicals, include synthetic rubber, plastics, textiles, fertilizers, insecticides and detergents.

Not all rubber is synthetic. Natural rubber is made from latex exuded by a rubber tree. Rubber trees, however, are tropical plants that do not thrive at all in Canada. It was a comparison of petroleum and natural rubber that gave the clue to making synthetic rubber. Both are made up of hydrocarbons; they are chemically related, though distantly. Nature arranged hydrogen and carbon in one combination to make petroleum, and in another to make rubber.

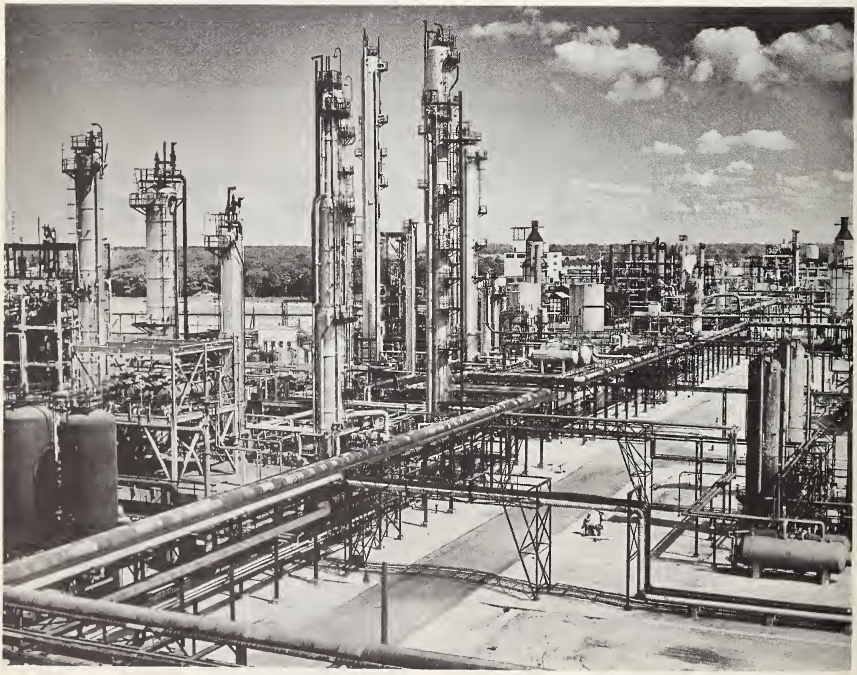
The basic unit in rubber is a substance called butadiene, which is



$$\begin{array}{ccccccc} H & & & & & & H \\ | & & & & & & | \\ C & = & C & - & C & = & C \\ | & & | & & | & & | \\ H & & H & & H & & H \end{array}$$

Rubber is made by linking butadiene molecules into long chains. So the real trick was to juggle the hydrocarbon atoms

Fig. 20-5. Part of a synthetic rubber plant. This shows the feed preparation area where the raw materials are prepared for the synthesis reaction.



in petroleum so that they became rubber. It took many years of research to learn how to do this but, eventually, chemists were able to convert butane (from petroleum or natural gas) into butadiene. Then they discovered that butadiene molecules could be linked into long chains in the presence of metallic sodium as a catalyst. These long chains are strong and flexible; they are a kind of rubber. This particular synthetic rubber is called Buna, a word in which "bu" stands for butadiene; "na," you will recall, is the symbol for sodium.

Coal, Petroleum and Natural Gas. The petroleum industry, once concerned entirely with producing gasoline, kerosene, fuel and lubricating oils, has, like the coal industry, turned to the production of chemicals. Nevertheless, coal, petroleum and natural gas are still important fuels and are likely to remain so until atomic energy becomes cheaper than the energy of coal, or until scientists learn how to trap and store energy from the sun efficiently and cheaply. In 1959 petroleum supplied 54.4 per cent of Canada's B.t.u., coal 20.8 per cent, and natural gas 14 per cent. The remaining 10.8 per cent was supplied by water power.

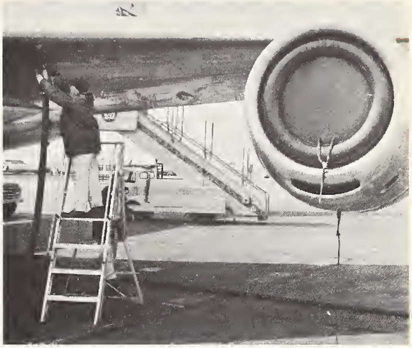


Fig. 20-6. Fueling a jet aircraft.

Things to Remember

Destructive distillation is the decomposition of a complex substance by heat, and a condensation of some of the vapors. Coke, tar, and coal gas are formed in the destructive distillation of coal.

Coal gas is a mixture of hydrogen, methane, and carbon monoxide.

Benzene and toluene are the most important liquid coal tar products. Benzene, C_6H_6, is the basic unit of a large number of hexagonal, ring-shaped molecules.

Crude Oil is separated into the following useful ***fractions*** by distillation: (1) gasoline, (2) kerosene, (3) fuel oil, (4) lubricating oil, (5) vaseline and paraffin wax.

Gasoline is a mixture of pentane, hexane, heptane, and octane.

Cracking is the process of breaking carbon chains.

Natural gas is a mixture mainly of methane and ethane.

Carbochemicals are chemicals from coal; ***petrochemicals*** are chemicals from petroleum.

Questions

GROUP A

1. Name the three products (solid, liquid and gas) obtained from the destructive distillation of coal.

2. Write the formulae and the structural formulae for (a) benzene, (b) toluene.
3. Name four of the ingredients of crude oil.
4. Name four of the constituents of gasoline and give their chemical formulae.
5. What is meant by cracking fuel oil? Why is this an important process?
6. What is natural gas? What is its heat value?
7. State the meaning of the terms petrochemicals and carbochemicals. Give examples.

GROUP B

8. What is the difference between distillation and destructive distillation? Explain.
9. Write balanced equations for the burning of the three main ingredients of coal gas.
10. Discuss the importance of coal by-products in chemical industries.
11. (a) Write the structural formula for a molecule of kerosene ($C_{10}H_{22}$).
    (b) Why is kerosene not a suitable fuel for the ordinary internal-combustion engine? Explain.
12. Write balanced equations for the complete combustion of (a) heptane, (b) hexane.
13. Explain how and why gasoline is used as a fuel in an internal-combustion engine. What causes the driving force?
14. Explain clearly why heat and pressure cause long-chain compounds to crack.
15. Write equations for the complete combustion of natural gas (methane, ethane, propane (C_3H_8), and butane (C_4H_{10})).
16. (a) What is butadiene? (b) Write its (i) formula, (ii) structural formula. (c) Why is butadiene an important compound in industry? Explain.
17. Draw the structural formulae of all the possible isomers of pentane.

Things to Do

Cracking oil · Crude oil (or medicinal mineral oil, which is a product of petroleum) can easily be cracked by heat and a catalyst. Pour 7 or 8 cc. of crude oil (or, better still, mineral oil which you can obtain at a drug store) into a one-inch test tube. Now pack the test tube, above the oil, with fine, loose steel wool. Place one end of a bent glass tube in a one-hole stopper and insert the stopper in the test tube. Force the other end of the bent tube through a two-hole stopper and into another one-inch test tube. When the two-hole stopper is inserted in this test tube the end of the glass tube should be about an inch and a half from the bottom of it. Bend another piece of glass tubing, an outlet tube for the escape of gases. One end of this tube should be drawn out to a jet (like the end of a medicine dropper). The arrangement is shown in Fig. 20-7.

Use two Bunsen burners for heating. Place one under the oil and heat it with a small flame. Hold the other in your hand and heat the lower part of the steel wool and the oil with a small flame. Gradually increase the heat of the burners until the oil boils. Now maintain the heat, and soon white fumes appear and escape at the jet. Ignite the fumes with a lighted taper—they burn with a bright flame.

As the oil molecules pass through the steel wool some of them are cracked to gasoline and other short-chain molecules which burn at the jet outlet.

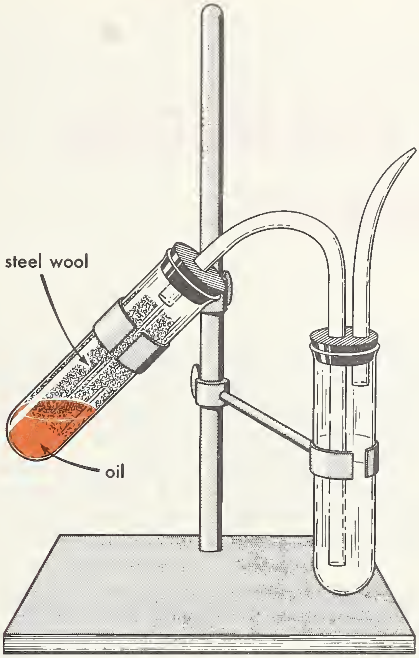


Fig. 20-7. Cracking oil.

21

SOME CHEMICAL INDUSTRIES

In 1867 the work of the Fathers of Confederation resulted in the birth of Canada. At that time in our history, chemical industries had hardly begun, and many of the products we now accept as essential to our way of life were unknown. Except for railroads, transportation was of the horse and buggy type—no automobiles or aircraft. Then again there were no plastics, no synthetic fabrics, no detergents, no insecticides, no antibiotics or artificial fertilizers.

These changes have come largely through chemical research, which is now an essential and important branch of industry. Every large chemical company expends from three to five per cent of its income on research. Chemical discoveries are rarely made by one man. Usually they are the result of the combined efforts of a team of researchers over a long period of time. When next you read that a new fabric or plastic has gone into production, you may be sure that the preliminary work was a joint effort of 10, 20, 30, or even more research chemists and technicians.

Chemical industries are really a combination of "farm" and factory. The "farm" supplies raw materials, and the factory converts these raw materials into finished products. Some countries, England and France for example, have highly organized factories but relatively poor "farms." As a result, they import a large proportion of their raw materials. Other countries, like Canada, the United States, and Russia are in a more favorable position. Their "farms" supply raw materials more abundantly, so that they are less dependent upon imports.

Chemical industries can be classified into three groups. First there are those which depend entirely upon nature for their raw materials. The paper and sugar industries are of this type. Both cellulose and sugar are products of photosynthesis, a process the chemist has not matched.

In the second group the chemist imitates and often improves upon the products of nature. These industries include rubber and fabrics. Rubber is a product of a tropical plant but the chemist can also make rubber from by-products of coal and oil. In fabrics, wool is an animal product and cotton a vegetable product. But again the chemist can make synthetic fabrics like nylon or rayon from the by-products of coal and oil.

Products in the third group are not reproduced in nature; they are entirely manmade. Plastics, explosives, and some of the new drugs are in this category.

The above is a varied list of industries—too extensive to be considered at length in this book. We will, however, make a selection from each of the three groups, beginning with the manufacture of paper.

PAPER

Since nature provides the raw material it is not surprising that papermaking is one of the oldest chemical arts. Papyrus, made from the pith of papyrus plants, was known to the Egyptians in 3000 B.C. It was probably the earliest writing material.

True paper is simply a sheet of matted cellulose fibers. This was first made in the second century A.D. However, it had no great impact on civilization until the middle of the 15th century. Then came the invention of movable type and Gutenberg's great masterpiece, the Mazarin Bible, was published in 1456. With the new emphasis on printing, the standard of literacy was raised throughout Europe. Indeed, some historians assert that the manufacture of paper made possible the revival of Europe and the Renaissance by stimulating the intellectual life of mankind.

A sheet of paper is composed of millions of tiny cellulose fibers matted together to a uniform thickness. Cellulose fibers may be derived from cotton or wood. The best writing papers are made from cotton rags. Pieces of cloth are torn to shreds, then to fibers, which are made into a homogeneous web of paper. Ordinary book paper is made from wood fibers, and this is the only process we shall discuss.

We must remember, however, that paper has many uses besides printing. We are familiar with paper towels, paper napkins, paper cartons, wrapping paper, and writing paper. The pulp and paper industry provides Canada with one of its major exports.

Paper from Wood. When trees are cut down the timber is aged for two or three years before it is used. The bark is then

Fig. 21-1. Overlooked by protective fire towers, stacks of pulpwood await processing at Hinton Mill, Alberta.

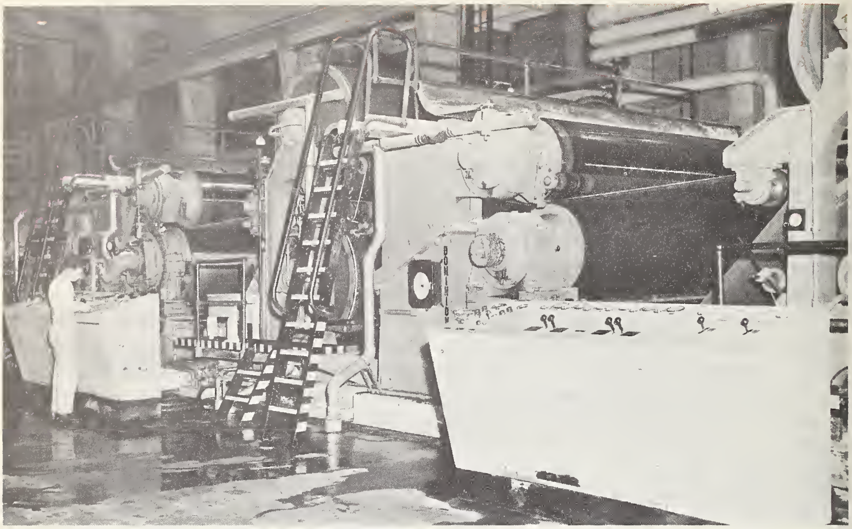


Fig. 21-2. This picture shows the rollers and driers used in making pulpwood fiber into sheets for shipment.

stripped off, and the wood ground to fine shreds and mixed with water to form wood pulp. The pulp is poured onto the screen of a paper machine, where it is dried to form paper.

This is the cheapest form of paper; it has little strength. Moreover, it contains objectionable substances like lignin and resins, and it slowly turns yellowish-brown when exposed to air. This particular paper is used for newspapers, where its poor quality is not a serious handicap.

For books, paper must be strong, white, and non-absorbent. To impart these qualities, the wood pulp must be treated chemically. First of all, lignin and resins are eliminated. If the wood is from evergreens (such as pine, spruce or hemlock), the gums and resins are best removed by calcium bisulfite. To do this, the shreds are cooked with high pressure steam and calcium bisulfite for several hours. The lignin, gums and resins dissolve in the bisulfite and are removed by repeatedly washing the pulp. The pulp is now bleached by chlorine, since even residual impurities in the pulp would impart a yellow color to the paper.

The pulp is then "beaten" for several hours. This means that the pulp is passed under a revolving drum set with "knives" which rub the fibers against other "knives" in a bedplate. This rubbing and cutting action separates the fibers and frays their ends so that they will adhere to make a strong paper. Sizing materials and fillers are then mixed with the pulp. The purpose of sizing is to make the paper impervious to water or ink. China clay or calcium sulfate (gypsum) are used as fillers to make the paper opaque and to give it a smooth surface.

The milky fluid from the vat is now ready for the papermaking machinery. This is a series of machines about 100 yards long and entirely automatic. It is one of the most efficient devices in chemical industry.

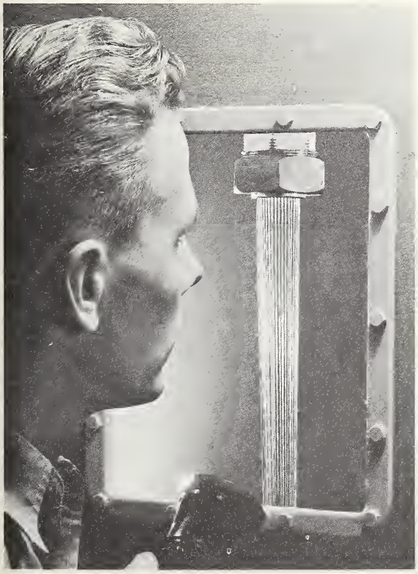
First the fluid is poured onto a moving wire screen which shakes sideways and moves forward at the same time. The motion shakes out much of the water and the cellulose fibers adhere to form a mat or web. The web then passes through a suction box where more water is removed. It then passes through rollers, where still more water is squeezed out. Finally, the mat passes over steam-heated rollers. The heat melts the sizing, which seeps into the paper, filling the pores. The finished paper emerges from the dry end of the machine and is wound into rolls.

TEXTILES

For thousands of years animal fibers (wool and silk) and vegetable fiber (cotton) have been used in making textiles. When we speak of fibers we mean the wool strands from a sheep's back or the filaments from a cotton boll. These fibers are cleaned and then, by twisting, spun into long threads called ***yarn.*** The yarn is then woven into cloth, or textile.

Until relatively recently wool and cotton were our only choices in selecting fabrics for clothing. Wool, which is warm and heavy, was usually selected for winter wear, and cotton, which is light, porous, and absorbent, was worn in summer. Special fabrics were also made of linen and silk.

Fig. 21-3. Acetate rayon yarn. Cellulose acetate is dissolved in acetone, and the solution is forced through a spinneret at the top of a shaft. The warm air in the shaft evaporates the acetone, and the fine streams of cellulose acetate coagulate into filaments which emerge from the shaft as a long strand of yarn.



Then in 1911, the first man-made fiber, ***rayon,*** was manufactured. Rayon, like cotton and linen, is almost pure cellulose; and, like paper, it is made from wood pulp and cotton linters. To make the fibers cellulose is dissolved in a suitable solvent. The fluid is then forced through fine pores in a nozzle, called a ***spinneret,*** to form a continuous filament which instantly solidifies. In one kind of rayon (viscose rayon) the filaments are formed by the coagulating action of an acid bath; in another (acetate rayon) they are formed by contact with air. Since the filament produced by a spinneret is continuous, the spinning of yarn is an unnecessary step in fashioning cloth.

Why have synthetic fibers become so important when nature provides wool and cotton so generously? A manufacturer can control the length and diameter and therefore the strength of synthetic fibers whereas the dimensions of cotton and wool fibers are fixed by nature. Moreover, at least 25 different kinds of synthetic fibers have been invented, all with different properties. Indeed, in the qualities of strength, wear, appearance, and resistance to shrinkage, synthetic fibers are superior to natural ones.

Rayon, as you know, is widely used in clothing, draperies, and upholstery but most of it is used in making tires. It is the cheapest of the synthetic fibers and by far the most plentiful.

Nylon. Rayon is not really a synthetic fiber because it is made from materials provided by nature. The first synthetic fiber, Nylon, was made by Du Pont Company in 1935. It has some remarkable properties. It is far stronger than any natural fiber and is much more resistant

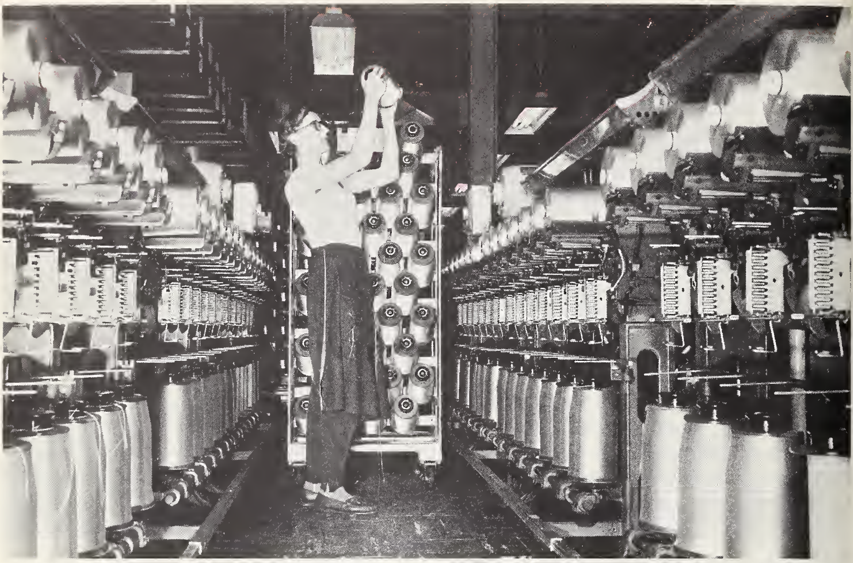
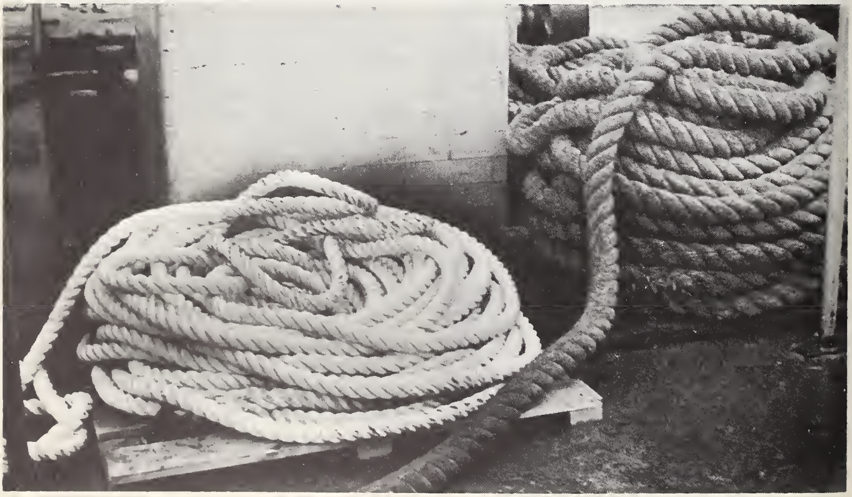


Fig. 21-4. Winding nylon on spools for shipment.

to water and chemicals. Moreover, the fibers are elastic and lustrous.

Nylon is formed by the reaction of two substances (adipic acid and hexamethylene) obtained from air, water, and coal or petroleum. The hot product, liquid nylon, is forced through the small openings of a spinneret and the fine streams

Fig. 21-5. The 5½-inch nylon rope (on the left) is as strong as the 8 inch natural fiber rope (on the right). Besides its great strength, nylon resists rot and mildew, and can be stored wet without damage.



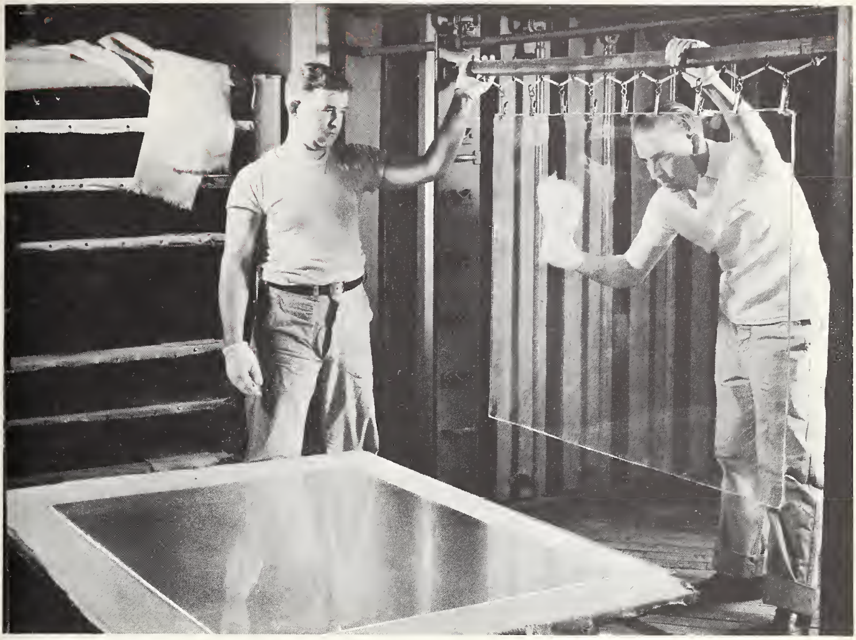


Fig. 21-6. Lucite sheeting. Notice its remarkable transparency.

are congealed at once by a blast of cold air. These solid threads are stretched to form the fibers used in weaving cloth or stockings.

Nylon is a versatile substance. It is used for many other purposes besides dresses and sweaters. Seat coverings, tire cords, hair brushes, racquet strings, and even rope indicate the scope of nylon products.

Today chemists are interested in nonwoven fabrics. Such fabrics are made by binding the fibers into a felt by heat or adhesives so that the processes of spinning and weaving will be eliminated. Felted cloth will be much cheaper than woven fabric and for some purposes just as attractive.

Plastics. The first plastic was made in 1868, by John W. Hyatt, in Albany, New York. He was competing for a prize offered for a material that could be substituted for expensive ivory in billiard balls. He experimented with nitrocellulose and camphor and, instead of ivory, he made a substance which he called ***celluloid.*** Celluloid is brittle, inflammable, and darkens with age so that its uses were limited. However, it found its way into men's wear as collars and cuffs, and the earliest roll film was made of it.

In 1909, Leo H. Baekeland, an immigrant Belgian chemist, experimented with phenol and formaldehyde hoping to discover a material that could be substituted for shellac. When he heated phenol and formaldehyde in alkaline solution a resin was formed which could be molded under heat and pressure into any desired shape. Baekeland called his product Bakelite. It is a ***plastic.***

The word ***plastic*** means a substance that can be softened by heat and then molded into shape. Dozens of plastics have now been made, but only two, Bakelite and Lucite, will be mentioned. Bakelite is a ***thermosetting*** plastic; that is to say, heat changes it to a permanently infusible solid. Lucite, on the other hand,

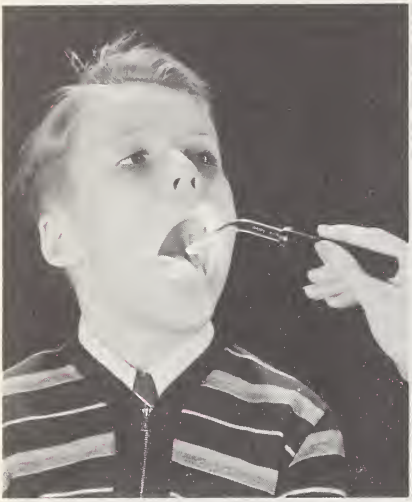


Fig. 21-7. This is a combination flashlight and tongue depressor made from lucite. The lucite rod transmits light from the flashlight bulb on the right, even though the rod is bent.

is ***thermoplastic;*** it can be repeatedly softened by heating.

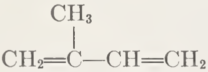
Lucite is a complex organic substance prepared from the raw materials coal, salt, water and petroleum or natural gas. It can be machined, sawed, and drilled like wood, and is the most transparent substance known. It is used for windows in airplanes, as safety "glass" in automobiles and as unbreakable lenses in spectacles.

Plastics are by far the most versatile of all synthetic substances. They are light, strong, easily fabricated, good electrical insulators, resistant to weather and most chemicals. These are valuable properties, and it is no wonder that their uses are legion. They are used in telephones, television cabinets, tableware, wall tile, gears for motors, weather balloons, food packages, electrical insulators, and bodies of sports cars. Many parts of an automobile are made of plastics—the steering wheel, the seat covers, knobs on the door handles. And how many others?

RUBBER

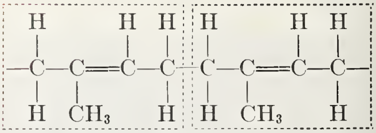
Natural rubber is obtained from a tropical plant cultivated chiefly in Malaya, Indo-China, and the East Indies. If the bark of a rubber plant is cut, a milky fluid called ***latex*** oozes out. Latex is an emulsion of rubber globules in water. If a dilute acid is added to it the rubber coagulates. The rubber is then pressed between rollers into thin sheets.

Natural rubber is a hydrocarbon; it contains a substance called isoprene, C_5H_8, or



$$CH_2{=}\overset{\displaystyle CH_3}{\overset{|}{C}}{-}CH{=}CH_2$$

Isoprene molecules have double bonds which enable them to link together to form enormously big molecules of rubber-like consistency. That is to say, isoprene in rubber unites in this way



$$-\underset{H}{\overset{H}{C}}-\underset{CH_3}{C}{=}\overset{H}{C}-\underset{H}{\overset{H}{C}}-\underset{H}{\overset{H}{C}}-\underset{CH_3}{C}{=}\overset{H}{C}-\underset{H}{\overset{H}{C}}-$$

Thus the secret of synthesizing rubber is to start with a simple hydrocarbon whose molecules have double bonds which make it possible for those molecules to unite to form large molecules. Chemists have now synthesized "rubber" from the by-products of coal and petroleum. As a result we are no longer dependent upon imports from tropical countries for the manufacture of this strategic material. At least four different kinds of rubber are manufactured but only one, Buna S (which was referred to in Chapter 20), will be discussed.

Buna S is the cheapest of synthetic

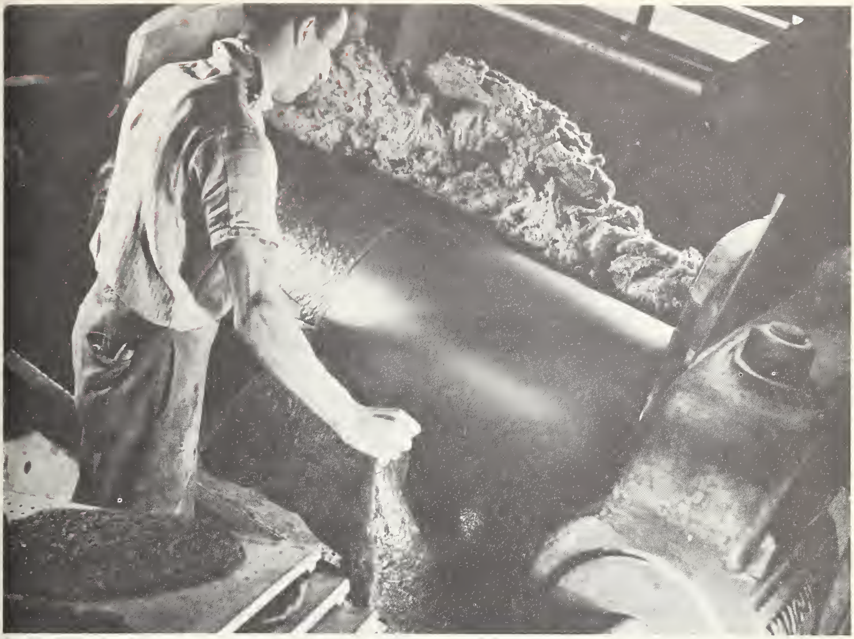
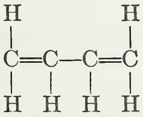


Fig. 21-8. The final process in the manufacture of synthetic rubber. Latex is fed through a slit onto the outside of a steam heated rotating drum. The heat softens the latex, and the film of rubber is stripped off the drum by the operator.

rubbers and it most nearly resembles natural rubber. It is manufactured by agitating a mixture of butadiene and styrene for several hours in soapy water. They combine to form a milky emulsion which resembles latex. As in the case of latex, the rubber is coagulated by acid and the product filtered and dried.

Both butadiene and styrene are hydrocarbons, obtained from petroleum, natural gas, or coal. Butadiene, a derivative of butane, is C_4H_6, or

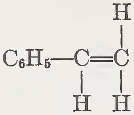


$$\begin{array}{ccccccc} \mathrm{H} & & & & & & \mathrm{H} \\ | & & & & & & | \\ \mathrm{C} & = & \mathrm{C} & - & \mathrm{C} & = & \mathrm{C} \\ | & & | & & | & & | \\ \mathrm{H} & & \mathrm{H} & & \mathrm{H} & & \mathrm{H} \end{array}$$

Styrene, a derivative of benzene, is a

Fig. 21-9. Synthetic rubber tires.

colorless liquid with a spicy odor. Its formula is $C_6H_5C_2H_3$ or



$$C_6H_5{-}\underset{H}{\overset{|}{C}}{=}\underset{H}{\overset{H}{C}}$$

The double bonds enable these two compounds to unite into large molecules with properties resembling rubber.

Vulcanization of Rubber. Natural rubber is sticky and soft at ordinary temperatures and hard and brittle at low temperatures. Moreover, rubber dissolves in gasoline and other organic solvents. These are undesirable properties and, unless corrected, would prevent the use of rubber in tires.

In 1840, Charles Goodyear, an American chemist, discovered that rubber can be hardened by heating latex with sulfur. The process is called ***vulcanization*** and it is an essential part of the manufacture of all rubbers. It is supposed that sulfur atoms attach themselves to the rubber molecules at the double bonds, thereby changing the properties of the molecules. Vulcanized rubber has great tensile strength, it is elastic and not easily dissolved by gasoline. It is, therefore, an ideal product for automobile tires. Rubber tires usually contain about 3 per cent of sulfur. If the sulfur is increased to 30 per cent "hard" rubber is formed. This was used to make combs and fountain pens before plastics took over.

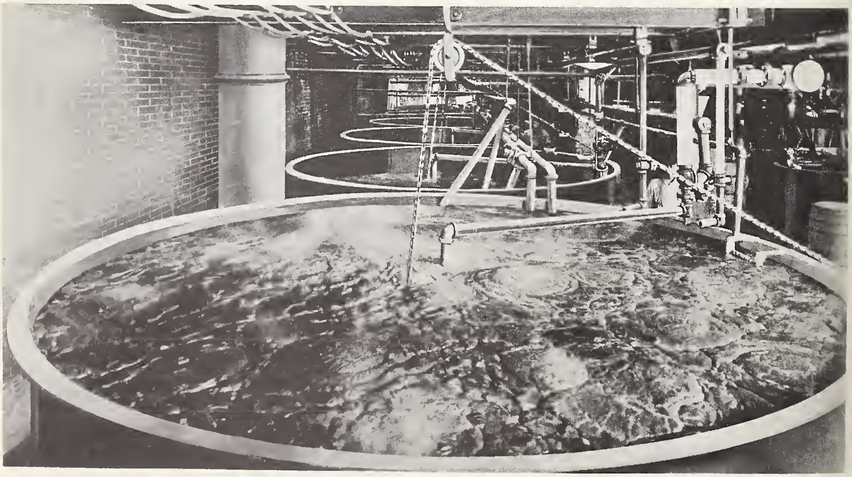
SOAP

It has been said, probably facetiously, that the civilization of a people can be measured by the amount of soap it uses. Certain it is that soap has been used as a cleansing agent since earliest times. Indeed, a well-equipped soap-making factory was discovered in the excavations of Pompeii.

It is only relatively recently, however, that the manufacture of soap has assumed the properties of a major industry.

In your great-grandfather's day, particu-

Fig. 21-10. This enormous soap kettle is about 30 feet in diameter and extends through 3 stories of the building.



larly if he lived on a farm, soap making was often a part of the household chores. Soap was made from animal fat and wood ashes. Wood ashes leached with water give a solution from which "lye" (potassium carbonate and potassium hydroxide) can be recovered. A wood fire was built under a large iron kettle which contained the mixture of lye and fat. The hot contents were stirred with a stick and a thick semifluid or soft soap was formed. The present day method of manufacturing soap is merely an elaboration of this simple procedure.

The Manufacture of Soap. Soap is formed when fats or oils are decomposed by sodium hydroxide or potassium hydroxide. Sodium hydroxide forms a hard soap; potassium hydroxide a soft soap such as shaving cream and liquid soap. Animal fats (beef fat, lard, or tallow) are suitable, or vegetable oils (cottonseed oil, coconut oil, or olive oil) are suitable. Usually a mixture of fats and oils is used.

Ordinary toilet and laundry soaps are made in enormous steel kettles, three stories high, which may contain 100 tons of soap.

A mixture of fats and oils is pumped into the kettle and the right amount of sodium hydroxide solution is added. The contents are then heated by steam which enters through the bottom of the kettle. The steam heats the mixture and also stirs it up so that the substances react more readily. During the heating, which usually takes about three or four hours, fatty acids are released from the fats and oils. The sodium hydroxide then reacts with the fatty acid to form a salt. ***The salt of a fatty acid is a soap*** and the chemical name for household soap is sodium stearate. Expressed in simple language, the reaction is

$$\text{fat} + \text{sodium hydroxide} \rightarrow \text{soap} + \text{glycerine.}$$

Glycerine is an important by-product in the soap industry. It is used for many purposes as we shall read in a later chapter.

To separate the soap from glycerine, salt is shoveled into the kettle. The soap, being insoluble in brine, rises to the top and is pumped out of the kettle. It is then washed to remove any remaining sodium hydroxide which, if left in the soap, would be harmful to skin and hair.

The soap is then dried by hot air and squeezed through a die to form a long bar. The bar is cut into cakes and the cakes stamped and wrapped for marketing.

There is a project on soap making at the end of the chapter.

GLASS

Glassmaking was one of the earliest chemical industries—the Egyptians made it from sand, lime, and soda 3000 years ago. Indeed, this old Egyptian glass has almost the same composition as the soda-lime glass made today.

Window glass was invented sometime during the Middle Ages. Even in those days glassblowers knew how to color glass with rich reds, yellows, greens, blues and purples, and the windows in some medieval cathedrals in Europe are priceless works of art.

From the 13th to the 18th century Venice was the center of the glass-making art. Venetian craftsmen blew glass in molds, annealed it, cut it, and decorated it with gold leaf. They were the first to make mirrors by depositing a thin coat of silver on the back of a sheet of glass.

Glassmaking is the oldest North American industry. It dates from the settlement at Jamestown, Virginia, 350 years ago. Just outside the stockade the colonists built a small furnace in which they made glass beads for bartering with Indians.

But glass, you will recall, was made in nature long before it was made by man. Natural glass is called obsidian. Obsidian is a volcanic rock which cools so rapidly

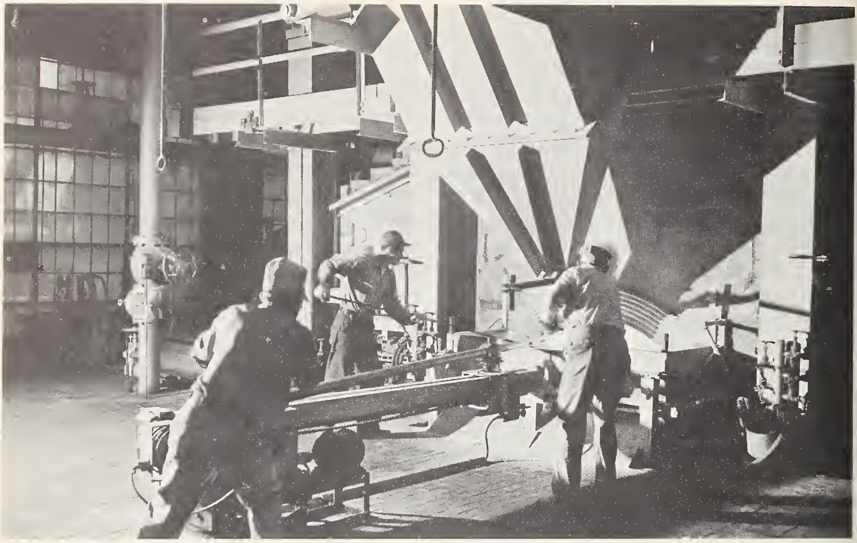


Fig. 21-11. The manufacture of glass. A charge of limestone, sodium carbonate, and silica flows down the shute (in the upper center) and into the "doghouse" at the mouth of the furnace. The two men near the furnace are pushing the charge into the furnace; the man in the foreground is distributing the charge in the furnace.

that large crystals were unable to form, a topic already discussed in Chapter 5.

The Manufacture of Glass. There are many kinds of glass, each with a specific use. Ordinary glass usually called ***soft or soda-lime glass,*** is made from a mixture of sand (SiO_2), limestone ($CaCO_3$), and sodium carbonate (Na_2CO_3). These raw materials are heated in a furnace. When they melt they react chemically as follows,

$$Na_2CO_3 + CaCO_3 + 4\ SiO_2 \rightarrow Na_2SiO_3 + CaSiO_3 + 2\ SiO_2 + 2\ CO_2$$

The molten glass, which is a mixture of sodium silicate, calcium silicate and silica, is then poured from the furnace and used for various products.

This transparent solid mixture has peculiar and useful properties. First, glass does not have a sharp melting point; it softens gradually when heated. That is why it can be blown, or bent, or shaped into any desired form. Another useful property is its extremely low solubility in water, and its inactivity with chemicals. That is why it is used almost universally as a container for solutions.

Kinds of Glass. By using various ingredients glass with special properties can be made for particular purposes. Glass can be made lighter than aluminum or heavier than steel; it can be blown, sawed, or even woven. It can be fabricated into an amazing variety of articles.

Plate Glass. Plate glass is used mostly in automobiles and store windows. To make it, molten glass from the furnace is poured on an iron table and spread to even thickness by a heavy roller. The sheet is then annealed; that is, cooled very slowly by passing it through an oven over 300 feet in length. The annealing process greatly increases the strength of the glass. The glass sheets are then ground with an abrasive and polished by rouge on revolving pads.

Cut Glass. Cut glass is made for table ware and other special articles. This glass

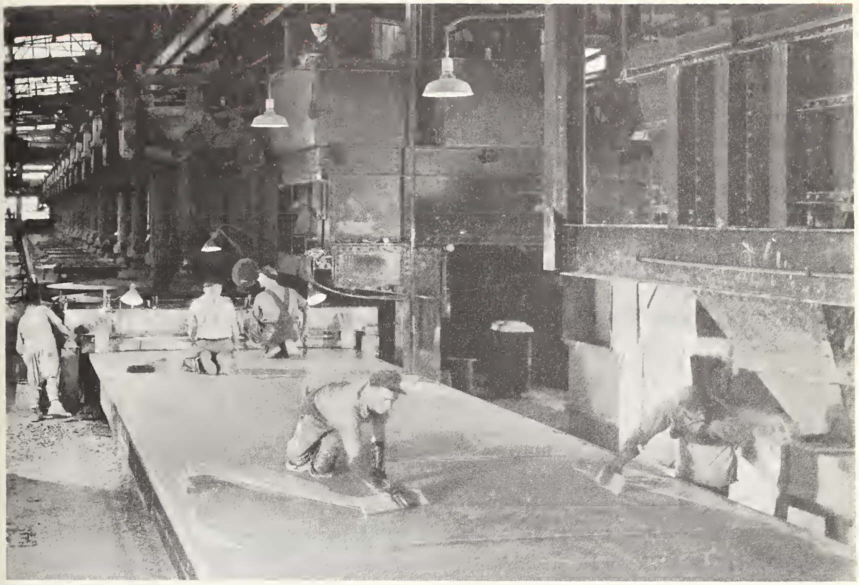
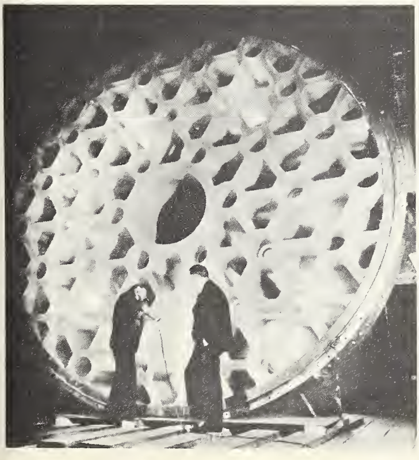


Fig. 21-12. Plate glass being prepared for grinding and polishing. The workmen are spreading plaster on the glass. The plaster cast keeps the glass perfectly flat as it is ground and polished on the large machines in the background.

contains silicates of lead and potassium, which affect the refractive properties of the glass, giving it an unusual sparkle.

Fig. 21-13. The underside of the 15-ton, 200-inch glass reflector of the Hale telescope.



Optical Glass. Optical glass is used for mirrors, lenses, and prisms, which, in turn, are used in optical instruments such as cameras, binoculars, telescopes, microscopes, and spectacles. This glass is similar to window glass except that it is made with particular care. Only ingredients of high-grade purity are used so that the glass is free from color, air bubbles, and other imperfections.

The reflecting mirror, a glass disk, in the Hale telescope on Mount Palomar is made of optical glass. It is the largest piece of glass ever made. The disk has a diameter of 200 inches and weighs about 15 tons. To reduce its weight and retain its strength, the back of the disk was made in the form of a grid. Eleven years of grinding and polishing were needed to make the surface optically perfect and, in the process, five tons of glass were removed.

Pyrex. Pyrex glass contains boron and aluminum as silicates as well as sodium and calcium. It has a higher melting point than ordinary window glass and also a smaller thermal expansion. It is therefore able to withstand sudden temperature changes without cracking. For this reason, it is used for baking dishes, skillets, coffee percolators and laboratory glassware.

Safety Glass. Safety glass is used for windshields and windows of automobiles. It is made by placing a layer of plastic between two sheets of plate glass. The three sheets are then firmly cemented by heat and pressure. If this glass is broken by a sharp blow the glass pieces adhere to the plastic so that danger of flying fragments is considerably reduced.

Glass Fibers. Finally, glass can be woven into cloth. Molten glass can be drawn through tiny holes to form a fluffy mass of fibers which can be used for insulating walls and roofs of buildings against heat losses. Glass fibers can also be drawn so fine that they can be twisted into yarn and woven into bedspreads, tablecloths, and fireproof curtains.

Things to Remember

Paper is simply a sheet of matted cellulose fibers; it is manufactured from wood.

Textiles are made from fibers.

Wool and cotton are ***natural fibers;*** rayon and nylon are made of ***synthetic fibers.***

Plastics are the most versatile synthetic substances made.

Natural rubber is made from the latex of a tropical plant; ***synthetic rubber*** is made from by-products of coal and oil.

Vulcanization is a process of hardening rubber with sulfur.

Soap is a salt of a fatty acid. The chemical name for common soap is sodium stearate.

Ordinary ***glass*** is made from sand, lime and soda (Na_2CO_3). It is called soft or soda-lime glass.

Questions

GROUP A

1. Chemical industries are a combination of "farm" and factory. What does this mean?
2. Classify chemical industries into three groups and give an example of each group.
3. Name two natural fibers and two synthetic fibers.
4. What are some of the advantages of synthetic fibers over natural fibers?
5. What is (*a*) a thermoplastic, (*b*) a thermosetting plastic?
6. State a similarity and a difference between natural and synthetic rubber.
7. What is meant by the terms (*a*) soft glass, (*b*) Pyrex, (*c*) safety glass?

GROUP B

8. Briefly describe how paper is made from wood.
9. How is rayon made? What is it used for?
10. What are plastics? Why has the manufacture of plastics become one of the largest chemical industries?

11. Describe the manufacture of soap.
12. What is meant by the term ***glass?*** Describe examples of (*a*) natural glass, (*b*) synthetic glass.

Things to Do

Soap making · To make a batch of soap we need fat or grease, and lye. The fat can be waste kitchen fat, or time can be saved by getting it at a meat store. If necessary, the fat (or grease) should be clarified by boiling it in twice its volume of water. This will also take salt out of fat (or grease), a necessary precaution. Why? Skim off about three pounds of clean molten fat and pour it in a dish, preferably an enamel utensil. While it is cooling prepare the lye solution.

Obtain a can of lye at a grocery store (the lye content is usually 12 or 13 oz.) Originally the word ***lye*** referred to the strong alkaline liquor of potassium carbonate made from wood ashes. Today it refers to any strong alkaline solution. The lye you purchase at the grocery store is called ***concentrated lye*** and it contains more than 90 per cent of solid sodium hydroxide.

Lye is dangerous to handle—it is poisonous, and it burns. Avoid getting it on your hands or particularly in your eyes. In case any lye should spill on your flesh, wash the affected parts with lots of water.

Pour a pint of cold water into an enamel dish and slowly add half the lye from a 12-oz. can. Stir the lye with a wooden rod until it has dissolved.

Now slowly pour the warm, not hot, lye solution into the warm grease and stir thoroughly for an hour or so.

A chemical change takes place during the stirring and the mixture begins to thicken. Now pour the creamy mixture into a rectangular cardboard box and let it stand for several days. The soap making goes on in the box; indeed, the longer reaction at the lower temperature in the box is precisely the same as the shorter reaction at a high temperature in the industrial process. When the soap has hardened, take it out of the box and cut it into cakes with a wire.

22

METALS

METALS ARE everywhere; we see them and use them every minute of the day. In the home we find aluminum or stainless steel cooking pans, copper wiring, steel radiators and furnaces, tungsten lamp filaments, silver cutlery, gold jewelry, zinc, copper and chromium in the refrigerator, and mercury in the thermometer. Outside, there is the same preponderance of metals, metals used for constructional work. Concrete structures have a framework of steel; even wooden houses are held together by steel nails. Locomotives are built of steel, airplanes of aluminum. All machines are made of metals. Have you ever thought of the metals that go into the construction of an automobile? The body, of course, is made of steel. But do you realize that zinc, copper, lead, tin, chromium, vanadium, manganese and many other metals are all a part of your automobile?

Metals are indispensable to our modern civilization, but it may surprise you to know that many metals were used by the Egyptians as early as 1000 B.C. Gold they found native—indeed, to them gold was cheaper than silver. Copper, iron, and tin they extracted from their ores. With these metals they fashioned daggers, vases, mirrors, and cooking utensils. They even made articles of bronze (or brass), an alloy of copper and tin. You will recall that St. Paul, writing to the Corinthians some 2000 years ago, says that without charity we "become as sounding brass or a tinkling cymbal."

Why is it that metals now play such an important part in our daily lives? First, let it be said that what we usually call metals are not metals but ***alloys***. Pure metals have remarkably few useful properties. Probably the only pure metals you have seen outside a laboratory are copper, mercury, and aluminum. Very few people have seen ***pure*** iron. It is soft and white like silver, and rusts easily. For constructional purposes it would be useless. In industry a metal must serve a specific purpose and its properties must be changed to serve that purpose. The metallurgist has made iron into dozens of different metals, or alloy steels, as we call them. By adding other metals to iron he can change its properties from soft to hard, from weak to tough, from brittle to malleable.

Metallurgists are constantly seeking new alloys, particularly high-temperature alloys. They have recently found a metal skin for missiles and supersonic planes, one that will retain its strength at the high temperatures caused by air friction.

Let us now consider the manufacture of a few metals. Metals, you will remember, usually occur in nature as minerals and the chemical problem in metallurgy is to extract the metal from its mineral. Not all minerals are suitable, however. They must be of sufficiently high grade to make it economically profitable to work them. Minerals from which metals are extracted are called ***ores***.

In this chapter we shall discuss only the three most important metals in industry and make brief reference to two that are now coming into production. Actually there are far more metals than nonmetals; about 70 of the 102 elements are metals. In short, metallurgy will receive scant treatment; for a more comprehensive picture, we must refer our readers to a chemistry book. Let us begin with iron and steel, because 90 per cent of the weight of all metal production is iron and steel.

IRON AND STEEL

Iron is widely distributed in nature; it comprises 4.5 per cent of the earth's crust. The most common ores of iron are black magnetite (Fe_3O_4) and reddish hematite (Fe_2O_3). Iron ore is found in many places in Canada: in British Columbia and Alberta, but especially at Steep Rock and Jamestown in Ontario, Wabana in Newfoundland, and in Labrador. Most of the Canadian iron is produced in Sydney, Nova Scotia, Hamilton and Sault Ste. Marie in Ontario.

A Blast Furnace. The metallurgical problem is to get iron out of an ore which contains Fe_2O_3 and a good deal of rocky material, chiefly silica. How can it be done? Actually, the oxygen is removed by coke and carbon monoxide, and the silica by limestone. This is done in a blast furnace, a huge steel tower. The furnace is lined with firebrick to withstand the intense heat. Hot compressed air is forced through openings (called tuyères and pronounced ***tweers***) at the base of the furnace, and waste gases are led off through a pipe near the top. Iron ore, limestone, and coke are continuously fed into the top of the furnace, and molten iron and slag collect at the bottom.

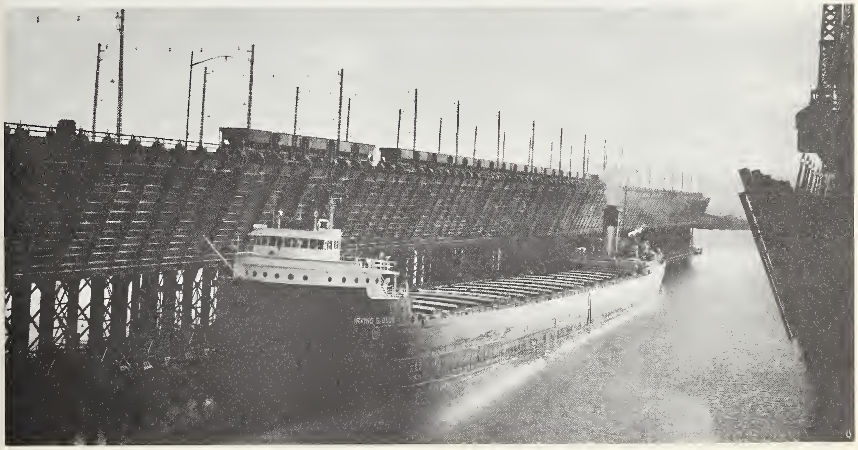
In the furnace, preheated air combines with coke to form carbon monoxide,

$$2C + O_2 \rightarrow 2CO$$

and the carbon monoxide and more coke then reduce iron oxide to iron,

$$\begin{aligned} Fe_2O_3 + 3C &\rightarrow 2Fe + 3CO \\ Fe_2O_3 + 3CO &\rightarrow 2Fe + 3CO_2 \end{aligned}$$

Fig. 22-1. This freighter has carried about 15,000 tons of iron ore across the Great Lakes.



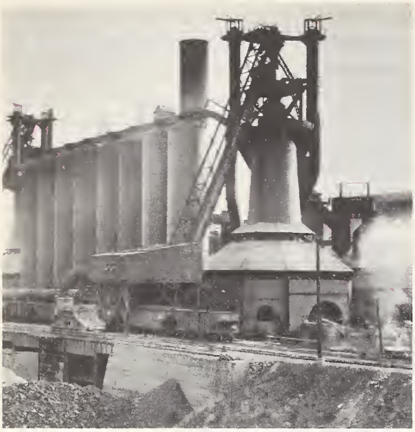


Fig. 22-2. A blast furnace in operation. On the left of the furnace are eight stoves for heating compressed air. A skip car carries raw materials to the top of the furnace up the steel inclined track shown in the picture. Notice that molten slag is running into a ladle on the right of the furnace.

Because heat is evolved in these reactions, and because the air is preheated, the temperature of the furnace is high enough to melt the iron. Molten iron sinks to the bottom of the furnace and is tapped off from time to time.

What happens to the silica? The silica reacts with the limestone to form calcium silicate, known as ***slag,***

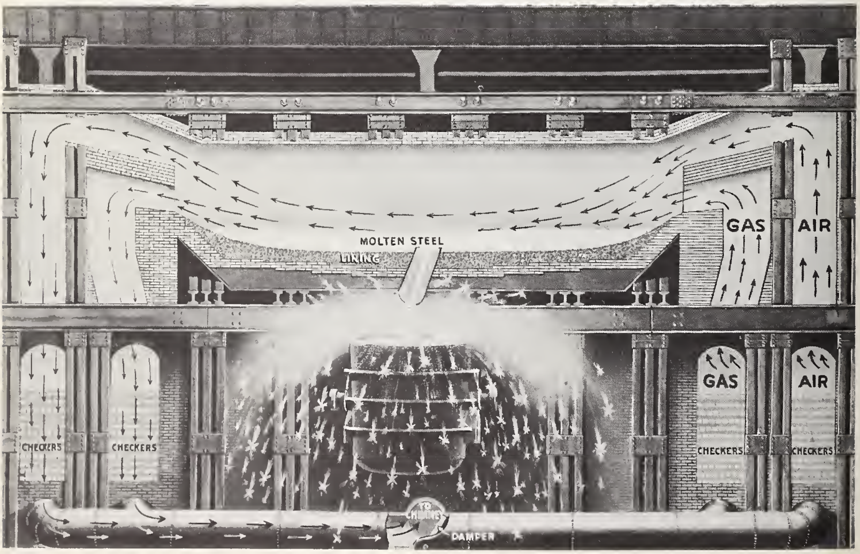
$$CaCO_3 + SiO_2 \rightarrow CaSiO_3 + CO_2$$

Slag has a much lower melting point than limestone and it also melts and falls, and floats on the molten iron. The slag is drawn off through an opening and hauled away to be used in making concrete for highway construction.

The molten iron, usually called ***pig iron,*** contains impurities such as carbon, sulfur, and phosphorus, which make the metal too brittle for constructional work. Some of it is used for making automobile parts such as cylinder blocks. Most of it, however, is carried by overhead crane to a nearby open-hearth furnace where it is made into steel.

An Open-Hearth Furnace. An open-hearth furnace is a shallow basin, lined

Fig. 22-3. An open-hearth furnace. Molten steel is heated by flames from the gas and hot air. The waste gases pass out through the checkers on the left and thence through the chimney.



with limestone, capable of holding more than 100 tons of liquid steel. At either side of the basin are two ducts, to admit fuel gas and air, and over it is a roof.

The furnace is charged with molten pig iron, scrap steel, and a little iron ore. Scrap steel comes from worn-out machinery, old automobiles, ships, bridges, and other similar sources.

Preheated fuel gas and air enter one side of the furnace. As they combine, long tongues of flame play over the charge which eventually melts completely. The waste gases escape by ducts on the other side heating up the checkerbrick inside the ducts. After 15 minutes or so fuel gas and air are led through the duct heated by the waste gases. This process is reversed every 15 minutes to insure that air and fuel gas are preheated. As a result, the temperature of the furnace is raised several hundred degrees, thus insuring that the whole charge is melted.

The iron ore supplies enough oxygen to oxidize impurities in the pig iron. Carbon burns to carbon dioxide, sulfur to sulfur dioxide and phosphorus to phosphoric oxide. Then various metals are added, depending on the particular kind of steel being made. All steels contain manganese and, in addition, they may contain one or more of the following metals: tungsten, chromium, nickel, vanadium, molybdenum. The whole operation takes about 12 hours. The molten steel is then poured into rectangular molds to form ingots which weigh about 2.5 tons. Everything in steel is made from ingots. The steel is stretched by squeezing a hot ingot between rollers. It is then run between other rollers which shape it for various industrial uses.

Steel. Steel is the strongest of all constructional materials. It consists chiefly of iron, a small amount of other metals, and carbon which may vary from 0.2 to 2 per cent. The greater the amount of carbon (up to 2 per cent) the harder is the steel. Hardness is not always desired, however; sometimes a softer steel is needed so that it can be bent and welded.

The uses of steel are legion. It is used in shipbuilding; in the manufacture of automobiles, locomotives, derricks, pipe-

Fig. 22-4. Stainless steel is a chromium-nickel alloy. It is used for fixtures and for utensils in kitchens.

lines, generators, and railroad tracks; in making cutlery, tin cans, and other articles.

Manganese hardens steel and is an ingredient in most steels. The metal tungsten also adds a special property to steel. Tungsten steel is used in making machine tools which must retain a sharp edge at high temperatures caused by cutting at high speed for long periods. Still another steel alloy is ***stainless*** steel which contains about 12 per cent of chromium and some nickel. It is used in high grade cutlery. And since it is not affected by acids or intense heat, stainless steel is used in constructing oil refineries and chemical plants.

Finally, let us return to that familiar article, a tin can. Its name is misleading because 98 per cent of a tin can is sheet steel. The steel, however, is plated with a thin layer of tin to protect it against the action of acids in fruits and vegetables.

ALUMINUM

Aluminum is a light, strong, metal; its density is only one-third the density of iron. Lightness is clearly a great advantage in constructional work. Moreover, aluminum is very plentiful—almost twice as plentiful as iron. It occurs in clay as alumina, Al_2O_3. One shovelful of the right kind of clay will yield a pound of aluminum. The high grade aluminum ore, which may contain as much as 50 per cent of the metal, is called ***bauxite.***

In spite of its abundance, aluminum was unknown until the 19th century and the aluminum industry did not get under way until about 1890. Why not? When alumina is heated with hot carbon nothing happens. That is to say the following reaction does ***not*** take place,

$$Al_2O_3 + 3C \rightarrow 2Al + 3CO$$

In other words, the bonds between aluminum and oxygen are stronger than the bonds between carbon and oxygen and so reduction of alumina by coke in a blast furnace is impossible.

Fig. 22-5. Unloading bauxite from ships at the docks at Port-Alfred, Quebec.

Aluminum from Bauxite. Charles Martin Hall, while still a college student, had the great idea that the bonds between aluminum and oxygen could be broken by electricity. For this process alumina must be in solution. After many failures Hall discovered that cryolite

Fig. 22-6. The Hall electrolytic process for manufacturing aluminum. Molten aluminum escapes through the outlet on the left and is collected in a ladle.

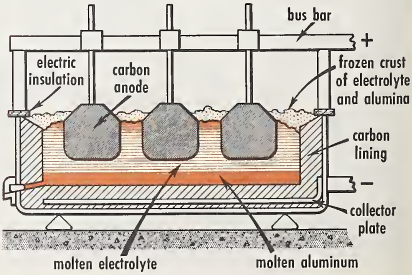


Fig. 22-7. This modern subway car is made of aluminum.

(Na_3AlF_6), a mineral found only in Greenland, melts at a low temperature and dissolves alumina easily. After this discovery his problem was relatively easy. He melted cryolite in a crucible, added alumina, and passed a current through the solution. Then, at the bottom of the crucible he found silvery buttons of aluminum. Thus began a mighty industry.

Even today aluminum is manufactured by the Hall process. Alumina is first separated from impurities in bauxite and then put in an electrolytic cell with cryolite. A heavy current enters the cell through carbon anodes, and leaves the cell via the carbon lining (called the cathode). The current melts the mixture and decomposes the alumina,

$$2Al_2O_3 \rightarrow 4Al + 3O_2$$

Molten aluminum collects at the bottom of the cell and is transferred to ladles and cast into ingots. Fortunately, manufacturers can make cryolite; they are not dependent upon imports from Greenland, although manufactured cryolite is considerably more expensive than the mineral.

Aluminum Metal. Aluminum is a soft silvery metal, not very strong. However, if it is alloyed with small amounts of copper, manganese, chromium and magnesium it becomes hard, strong and tough. It is the combination of lightness and strength that makes aluminum alloys so valuable in airplane construction. Aluminum alloys are also used to construct railroad cars, buses, bridges, ladders, chairs, lawn mowers, and a host of other articles including pots and pans.

Pots and Pans. Aluminum cooking utensils are in common usage in the average household. They have some advantages over utensils of copper and steel. However, it is unfortunate that an aluminum pan darkens if vegetables containing iron are cooked in it. The dark deposit is iron and is quite harmless in spite of the cry of "aluminum poisoning" sometimes heard from apprehensive housewives. The dark stain inside the pan will disappear if acid foods such as tomatoes or apples are cooked in it. This is because the acid reacts with the iron which goes back into the food.

COPPER

Copper has been known since ancient times. It was found native on the island of Cyprus, and the early Egyptians used it for decorative purposes. Combined with tin it forms a hard and durable alloy called bronze. Caesar's legions used knives, hatchets, and spearheads made of bronze. When St. Paul wrote of "sounding brass" he probably referred to what is now called bronze. Today brass refers to an alloy of copper, tin and zinc, and zinc is a relatively recent metal.

The invention of Edison's electric light bulb in 1878 was really the beginning of the large-scale manufacture of copper. Copper is a good conductor of electricity and this is its main use. An electric current from a generator is conducted by heavy copper wire to a transformer inside which is more copper wire. From the transformer electricity flows through a copper transmission line, often hundreds of miles long. From the transmission line electricity is taken into the home. Many millions of homes and other buildings are wired for electricity today. No wonder the electrical industry uses more copper than all other industries together.

Fig. 22-8. A copper miner drilling holes for dynamite before blasting the vein of copper ore.

Copper Metal from Copper Ore. Copper ores are unusually colorful: malachite, $CuCO_3 \cdot Cu(OH)_2$, is green; cuprite, Cu_2O, is red; chalcocite, Cu_2S, is multicolored. Yet, in spite of their color display, they contain very little copper. A high-grade ore of copper contains only about 5 per cent of the metal. This seems very little when compared with a good hematite ore, which contains 60 per cent of iron. Indeed, the recovery of this small percentage of copper from its rocky matrix is a most complex affair involving five different steps. After the fourth step molten copper is obtained from a furnace and cast into slabs. This copper contains small amounts of other metals (about one per cent in all) which would reduce its effectiveness as an electrical conductor. The final process is therefore the refining of copper by electrolysis.

To do this a slab of impure copper is placed in copper sulfate solution and a thin sheet of ***pure*** copper is also suspended in the solution. A current then flows into the slab (the anode) through the solution, and out at the thin sheet (the cathode). In this way, pure copper is carried from the anode and deposited on the cathode. This method of depositing copper can be shown in the following experiment.

Demonstration

22-1. Electrolytic deposition of copper

Suspend a copper sheet(about 3 in. by 4 in.) in a copper sulfate solution. Also place a stick of carbon (taken from a dry cell) in the solution. Arrange five dry cells in series. Now attach the positive terminal

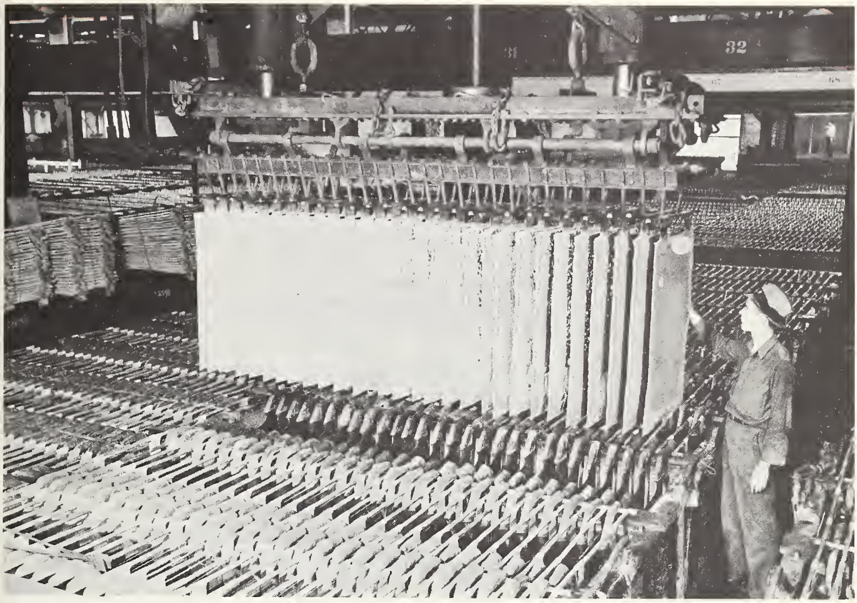
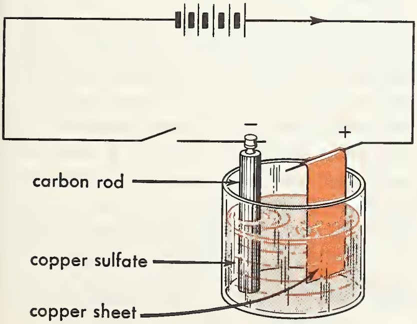


Fig. 22-9. Refining copper by electrolysis. A block of 30 copper cathodes have been lifted out of a tank containing copper sulfate solution. These cathode sheets are pure copper.

of the battery to the copper sheet and the negative terminal to the carbon rod. Close the switch and let the current flow for two or three minutes. Open the switch and examine the carbon rod. Notice the red deposit of copper that has plated upon it.

Uses of Copper. Copper wire is made from rods by drawing them through dies; diamond dies are used for making fine wire. For other purposes, however, pure copper is too soft and is hardened by making alloys. Bronze is an alloy of copper and tin; brass, a tough alloy of copper, tin and zinc. Monel metal, an alloy of copper and nickel, is the toughest of all copper alloys. It is as strong as steel and exceedingly resistant to corrosion.

Fig. 22-10. Electrolytic deposition of copper.



Copper and its alloys are widely used in constructing all kinds of vehicles, particularly ships. Marine fittings must be resistant to sea water, and that is why copper is used.

Gold and silver coinage are also alloyed with copper. "Sterling" silver (British silver) is an alloy of 92.5 per cent silver and 7.5 per cent copper. What we call "silver" coins in Canada contain 20 per cent of copper, the rest is silver. The amount of gold in an alloy is expressed in ***carats***. Pure gold is 24 carats. American gold coinage is 21.6 carats. Hence, as a percentage, the gold

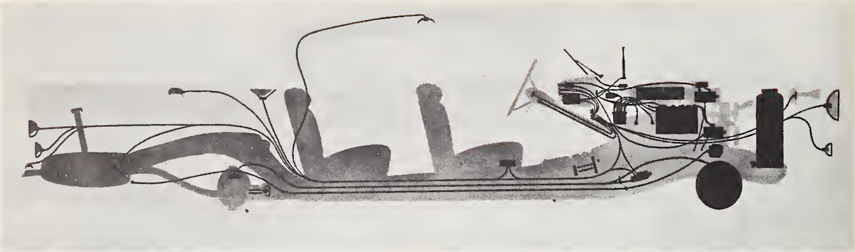


Fig. 22-11. The black parts show the many uses of copper and its alloys in an automobile. Besides its use in wiring, copper is also used in the brake fluid tubing, the brake cylinders, the radiator, crankshaft, carburetor, generator, and other parts.

content is $\frac{21.6}{24} \times 100$ or 90 per cent, and copper 10 per cent.

A Shortage of Metals. Metals are unevenly distributed in the earth's crust; no country has a monopoly of them, and all countries are short of some of them. And let us remember that minerals cannot be renewed—once they are taken from the ground there is no second crop of them. Even Canada, a favored country in mineral resources, is far from being self-sufficient. We do have many mineral reserves, especially in the north. These northern resources have been only partially exploited because of transportation difficulties. The new Great Slave Lake Railway to Hay River and Pine Point in the Northwest Territories will open up vast new reserves. Canada produces copper, nickel, silver, gold, zinc, lead, cobalt, and uranium, as well as lesser amounts of many other metals. But we must also import some metal ores, such as bauxite, which is used in the manufacture of aluminum, and chromite for the manufacture of chromium. Already, throughout the world, there is scarcity of metals. Can this problem be solved?

It is possible to make much greater use of aluminum and magnesium, both of which are abundant in the earth's crust. Then again sea water is an enormous reservoir of magnesium and other metals. But metals are so thinly scattered in sea water that it is exceedingly expensive to recover them. For example, 800 tons of sea water must be processed to obtain one ton of magnesium. The most promising future supply is to develop "new" metals such as vanadium, beryllium, titanium, and zirconium.

Titanium and Zirconium. Although titanium was discovered 160 years ago, it has been merely a laboratory curiosity until relatively recently. This is because the peculiar properties of the molten metal prevented the extraction of it on a large scale. For instance, molten titanium absorbs large quantities of oxygen and nitrogen which make the metal brittle. Therefore it is impossible to reduce the oxide ore in air. Instead it is melted in an inert gas such as argon or helium. Another difficulty is that the molten metal dissolves ordinary furnace linings and crucibles. In spite of these difficulties titanium is now being manufactured, and in a few years its production will reach industrial proportions.

Titanium is a most promising new metal. It is lighter than steel and, when alloyed with iron, chromium, and molybdenum, it is quite as strong as stainless steel and even more resistant to corrosion. Titanium is clearly a constructional material with a future. It is the fourth most plentiful metal, and fortunately there are reserves of it in Canada.

Zirconium, like titanium, was also discovered about 160 years ago. It, too, when molten, has a strong affinity for oxygen and nitrogen, and for refractories (furnace linings). Zirconium is extraordinarily resistant to all forms of corrosion—it is not attacked by air, sea water, and only very slightly by acids and alkalis. Nor does it readily deteriorate (as other metals do) when bombarded by radioactive radiations. For this reason, it is being increasingly used in the construction of atomic reactors.

Zirconium is expensive to manufacture. It is not as plentiful as titanium but far more abundant than copper, lead, or zinc.

Things to Remember

Ores are minerals from which metals are extracted.

Metals are usually used as ***alloys.***

In a ***blast furnace*** iron ore is made into pig iron.

In an ***open-hearth furnace*** pig iron is made into steel.

Charles Martin Hall discovered the modern method of extracting aluminum from bauxite.

Aluminum is a light metal, made by an electrical process.

Copper is used as an electrical conductor.

Brass and bronze are copper alloys.

Titanium and zirconium are "new" metals, both corrosion-resistant even at high temperatures.

Questions

GROUP A

1. Name three pure metals you have seen outside a laboratory and three you have seen only in alloys.
2. What is an ore? Name four ores of different metals.
3. What is pig iron? Why is it not used in constructional work?
4. Name four alloys of copper and, in each case, state the alloying metal or metals.
5. (*a*) What is the gold content of 20.0 carat gold? (*b*) What is probably the other metal present?

GROUP B

6. Describe briefly how pig iron is obtained from iron ore in a blast furnace.
7. What is steel? Discuss the particular properties and uses of two alloy steels.
8. How is the metal aluminum manufactured?
9. How are fuel gas and air preheated in the open-hearth operation? Explain why this precaution is necessary.
10. How is pure copper obtained electrolytically? Explain in detail.
11. Write balanced equations for: (*a*) the reduction of iron ore in a blast furnace, (*b*) the formation of slag in a blast furnace, (*c*) the preparation of aluminum from alumina.
12. What are some of the useful properties of titanium? Why is titanium a relatively "new" metal? Explain.

Things to Do

To make a mercury alloy · Mercury makes alloys with most metals very easily at ordinary temperatures. Mercury alloys are called amalgams. The dentist makes amalgams of mercury and silver or mercury and gold for filling teeth. These amalgams are soft when made so that they fill a cavity, and they harden in a few minutes.

An interesting way to form an amalgam is to dip a piece of copper into mercury. First the copper should be scrubbed with soapy water to remove surface grease. Then pour a little mercury into an evaporating dish and place the copper on the surface of the mercury. Notice that the copper floats. Then hold it under mercury for a few seconds with a glass rod. Remove the copper from the mercury and rub it with a cloth. What has happened to cause a color change? Remember also that gold and mercury amalgamate instantly. If you are wearing a gold watch or ring when you do this experiment, take precautions to prevent contact between the gold and mercury. Caution: mercury vapor is poisonous.

23

FOODS, DRUGS, AND ANTISEPTICS

During the past 60 years chemists, biologists and medical men have combined forces and produced an amazing array of chemotherapeutic agents—antiseptics, antitoxins, vitamins, and antibiotics. But nothing has been more rewarding than the joint research in the nutritional field. Until relatively recently this was a neglected field—the terms ***balanced diet*** and ***food value*** were unknown. It used to be thought that proper nourishment was merely a question of getting enough to eat. Today we know it is possible to starve in the midst of plenty. Deficiency diseases, for example, which may affect bones, skin and teeth are caused by something ***not*** eaten, something that is omitted from the diet.

In the days of sailing ships, sailors, during long sea voyages, were often stricken with scurvy, a disease that caused loosening of teeth, bleeding gums and soreness of joints. Then, in the 18th century, it was discovered that citrus fruits, such as limes, greatly reduced the incidence of scurvy. The British Navy thereupon supplied lime juice to its sailors. Science had not advanced far enough to take advantage of this important clue. Instead, the term "Limey" was applied to British sailors as a mark of scorn because they drank lime juice instead of rum.

The first significant step in understanding human nutrition came at the beginning of the present century. A study was made of the cause of the high incidence of beriberi amongst the natives of the Dutch East Indies. Beriberi is a debilitating disease that causes stunted growth, skin infection and inflammation of the eyes. It was discovered that the disease was the result of eating polished rice, the staple food of the natives. In other words, something of vital nutritional value was lost when husks were removed from rice. Then began the search which led to the discovery of vitamins, the first of which, vitamin A, was isolated in 1914.

FOOD

The body is like an internal-combustion engine. It "burns" food and releases enough energy to keep the body warm and to enable us to do work. But food does far more than supply energy—it is needed to form tissue. No child could grow without food to supply this living material. Even when we are fully grown, food is needed for this same purpose. Throughout life cells of the body wear away and disintegrate, and food must make good these losses in living tissue. When food enters the stomach and intestines it is attacked by "juices" and enzymes, and decomposed to smaller molecules—a process called ***digestion.*** The food particles are then small enough to pass through the intestine wall into the blood stream. They are then carried to the cells, where the oxidation reactions and repair work take place.

To supply the various needs of the body a "balanced diet" is essential. This should include carbohydrates, fats, proteins, minerals and vitamins.

Carbohydrates. What are carbohydrates? In food they are chiefly sugar ($C_{12}H_{22}O_{11}$) and starch ($C_6H_{10}O_5$). They are made by plants and are our most abundant source of energy. Cereals, rice and potatoes all contain a good deal of starch and bread, the "staff of life," has more than 50 per cent of it.

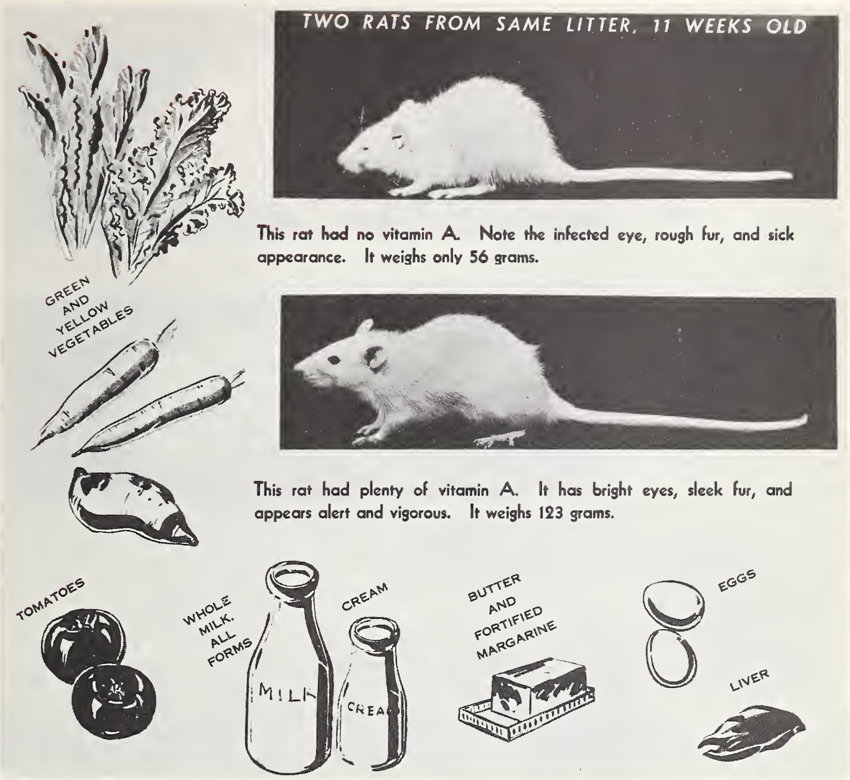
Proteins. Proteins contain nitrogen as well as carbon, hydrogen and oxygen. They are essential constituents of the cells of animals and plants. For this reason, much of our food consists of the tissues of other organisms. Beef, mutton, pork and fish are good sources of proteins. Meats and fish are considerably more expensive than carbohydrate foods. Proteins, however, occur in some vegetables such as peas, beans, peanuts, and soybeans. Eggs are also rich in protein.

Fig. 23-1. These foods are rich in proteins.

Fig. 23-2. These foods are rich in minerals and vitamins.

Fats. Fats are compounds of carbon, hydrogen, and oxygen, but they have a higher percentage of carbon than carbohydrates and they are therefore regarded as concentrated "fuels." Fats occur in animals and plants. Bacon, fat meats, butter and cream are all good animal sources of fat; many vegetable products, such as margarine, olive oil, corn oil, and cottonseed oil also contain it.

Minerals. Minerals are needed to build up tissue and for various body functions. A balanced diet must therefore contain small amounts of calcium, phosphorus, potassium, magnesium, iron, sodium, and sulfur. Calcium and phosphorus, for example, are essential in making bones and


TWO RATS FROM SAME LITTER, 11 WEEKS OLD
GREEN AND YELLOW VEGETABLES
TOMATOES
WHOLE MILK, ALL FORMS
MILK
CREAM
CREAM
BUTTER AND FORTIFIED MARGARINE
EGGS
LIVER

This rat had no vitamin A. Note the infected eye, rough fur, and sick appearance. It weighs only 56 grams.

This rat had plenty of vitamin A. It has bright eyes, sleek fur, and appears alert and vigorous. It weighs 123 grams.

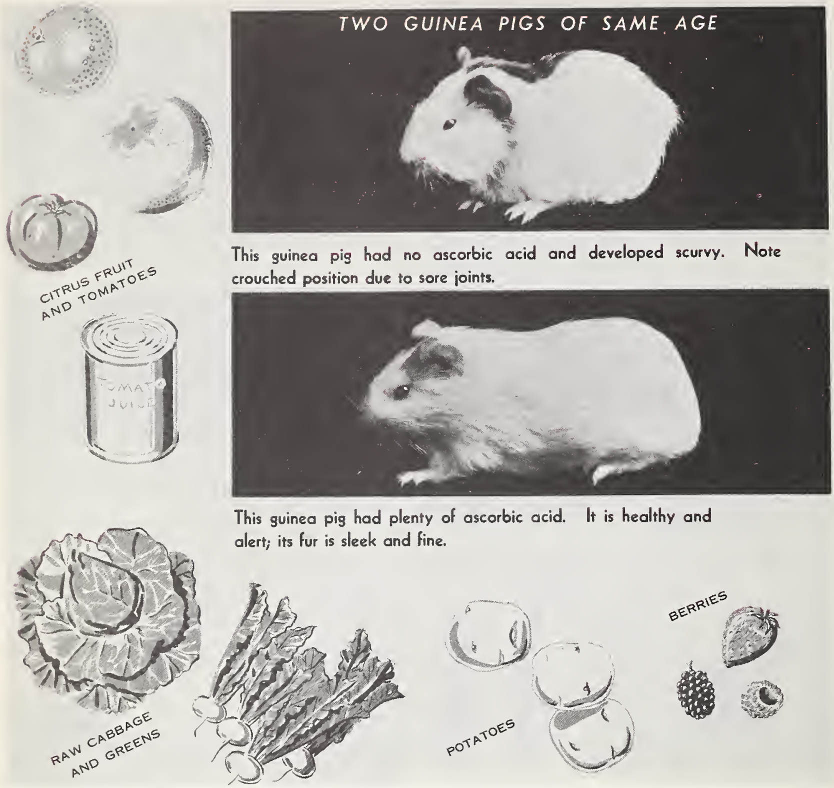
Fig. 23-3. Vitamin A is an essential ingredient in the diet.

teeth; iron is a necessary ingredient of blood.

All green vegetables contain minerals. That is why green vegetables should be a part of the daily diet. In foods such as rice, wheat, potatoes, and apples the minerals are concentrated in the outer layers. As a result, apples and potatoes lose their minerals if they are peeled. The same is true of white flour and polished rice, since the outer layers of wheat and rice grains are lost in the processing.

Vitamins. Vitamins are a ***vital*** part of the diet; they occur in minute quantities in many natural foods. Vitamins can be regarded as body regulators; without them various chemical changes in the body cannot take place, and deficiency diseases result. When first isolated their chemical composition was unknown and letters of the alphabet were used to distinguish them.

Vitamin A. Vitamin A ($C_{20}H_{29}OH$) was first obtained from butter in 1914. It is a pale yellow viscous liquid, insoluble in water but soluble in fats. Lack of this vitamin results in slow growth (poor bones and teeth), night blindness, and nervous irritability. Good sources of vitamin A are egg-yolk, fish-liver oils, butter, and fresh vegetables, particularly carrots. Fresh vegetables contain a yellow substance ***carotene*** which the body changes into vitamin A.



This guinea pig had no ascorbic acid and developed scurvy. Note crouched position due to sore joints.

This guinea pig had plenty of ascorbic acid. It is healthy and alert; its fur is sleek and fine.

Fig. 23-4. A comparison of these guinea pigs shows the need for vitamin C.

Vitamin B. Vitamin B was the second vitamin to be isolated. It has been found to be a mixture of at least three organic compounds, thiamin, riboflavin, and niacin. For this reason it is usually called vitamin-B complex. Milk, eggs, and fresh vegetables contain niacin and riboflavin. The richest sources of thiamin are brewers' yeast and whole grain foods such as wheat and rice. It was the removal of rice husks with its loss of thiamin that caused beriberi in the East Indians.

Vitamin C. Vitamin C is a white, crystalline, water-soluble solid known as ascorbic acid ($C_6H_8O_6$). Lack of it causes scurvy. Vitamin C is supplied by citrus fruits and by the vegetables lettuce, cabbage, and tomatoes. This vitamin is decomposed, however, by heat and by atmospheric oxygen so that only fresh raw fruits and vegetables are a good source of it.

Vitamin D. A deficiency of vitamin D prevents the proper absorption of calcium and phosphorus in the bones and is the cause of rickets. This vitamin occurs in milk, eggs and butter, but various fish liver oils are a rich source of it. It is sometimes called the "sunshine vitamin" because ultraviolet light changes cholesterol, a chemical in the body, to vitamin D.

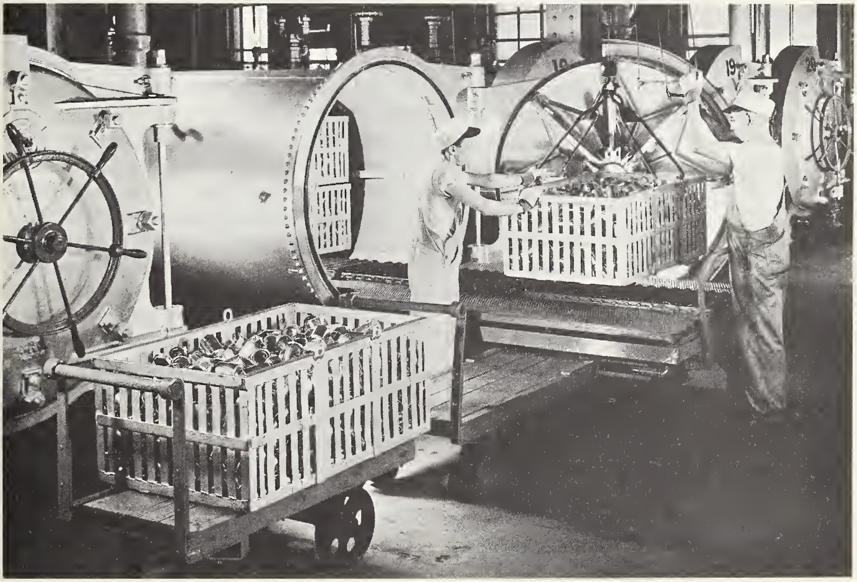
Synthetic Vitamins. Chemists have learned how to take matter apart systematically in order to find out how it was put together in the first place. This trick is the foundation of synthetic chemistry. All the vitamins (except D) have been taken apart, the pieces have been identified, and then put together again. In other words, vitamins A, B, and C have been synthesized and are now manufactured. This means that refined foods like white flour and polished rice can be enriched with vitamin B. Moreover, if vitamins are prescribed by a physician, vitamin pills can be purchased at reasonable cost.

PRESERVATION OF FOOD

The atmosphere contains myriads of microscopic plants called yeasts and bacteria. Neither yeast nor bacteria contain chlorophyll and, as a result, they are unable to manufacture food out of the carbon dioxide and water vapor in the air. Instead, they attack, decompose, and thrive upon organic material. Bacteria serve a useful purpose by decomposing organic material in soil, thus making it available as food for growing plants. But yeasts, and bacteria also attack foods such as meat, fish, eggs, milk, vegetables and fruit. The foods are decomposed, develop bad odors, and may even become poisonous. How can foods be preserved against the action of bacteria? This may be done by canning, by refrigeration, by preservatives, and by irradiation.

Canning. When fruits are plentiful they are cheap. This is the time to preserve them for use in winter and spring. In canning, all bacteria in the fruit are first destroyed by heat and then precautions are taken to keep air out of the jars. Why?

Fig. 23-5. A step in the canning process. The tin cans shown in the picture contain soup. They are all sealed. The basket of cans which is being lifted from the carrier to the rollers will be pushed into the giant retort. The retort will then be closed, and the cans heated to destroy all bacteria in the soup.



Jars are first sterilized in hot water and the fruit carefully selected. Fruit should not be bruised, otherwise bacteria may enter through the skin. Needless to say, hands should be scrupulously clean.

Fruit is packed into jars, which are then filled with syrup. The jars are placed in a boiler and almost covered with water. Rubber rings are fitted onto the jars and glass covers put loosely in place. The water is then strongly heated, enough to destroy all bacteria, and the cover is tightly sealed while the jar is still hot. If the jars are stored in a cool dark place, the fruit can be preserved indefinitely.

Refrigeration. Bacteria do not develop at low temperatures so that meat, fish, fruit and vegetables can be preserved in cold storage. Fish is best stored at 5° F, meat and poultry at 10° F. At these temperatures they can be preserved for months. Fruits should be kept at a higher temperature, just above the freezing point of water. A low temperature merely retards the ripening process and, when fruits are taken out of the refrigerator, ripening proceeds at a normal rate.

Pasteurization of Milk. Milk is an almost perfect food; it contains all the essentials of a complete diet—protein, fat, carbohydrate (sugar), calcium salts and other minerals, and some vitamins. It can be fermented in various ways to give more concentrated foods. Cheese, for example, is made this way and contains all the proteins and fats of the original milk.

Because milk is such an ideal food many kinds of bacteria thrive in it—typhoid, diphtheria, dysentery, and tuberculosis germs have all been found in milk. Milk should therefore be pasteurized as a safety precaution against harmful bacteria. In this process milk is heated to 145° F for half an hour and then quickly cooled. All injurious germs are thereby destroyed and the less harmful are rendered inactive so that milk remains fresh for 48 hours or so. It should be borne in mind, however, that pasteurization destroys some vitamins but these can be replaced by including orange juice in the diet.

Preservatives. It has been known for centuries that certain preservatives prevent the decomposition of food. There are numerous references in the Bible to the use of salt and spices for this purpose. Moreover, you will recall that Columbus tried to find a safer route to the Indies, which supplied spices to the Mediterranean people—spices to preserve their meat.

Preservatives destroy bacteria. In recent years various chemicals have been used as preservatives, some of them harmful. That is to say, in destroying bacteria they also adulterate food. The Food and Drugs Act is intended to protect the consumer against adulterated foods. Such preservatives as formaldehyde and boric acid are now forbidden by law, and less harmful ones like sodium benzoate may be used in only limited amounts. Whenever a chemical preservative is used, it must be so stated on the label of the container.

A concentrated sugar solution also inhibits bacteria. We are familiar with the use of sugar as a preservative in jams and jellies. Paraffin wax poured on the top of home-made jam excludes the air. Why is this necessary? Still another harmless and effective preservative is vinegar.

Irradiation of Foods. Gamma rays emitted by radioactive substances also destroy bacteria. This simple fact promises to revolutionize methods of preserving foods. If meat, fish, and vegetables are irradiated by various radioactive substances made in a reactor (see Chapter 33) they can be kept in fresh condition indefinitely without the need of refrigeration. However, food preservation by radioactivity is still in the experimental stage. Two important questions must be unequivocally answered before large-scale operations can begin: Do gamma rays affect the quality of foods, or make food toxic? Tests have been underway for some years; however, the questions have not yet been completely answered.

DRUGS

Drugs have been known since earliest times. Witch doctors of primitive civilizations learned by trial and error of the beneficial or harmful effects of concoctions of leaves, seeds, berries, roots, and barks. The two well-known drugs, cocaine and quinine, were known to the Indians of Peru in the 17th century. They discovered that pain was relieved by chewing coca leaves, which we now know contain the habit-forming drug cocaine. But it was not until about 1890 that cocaine came into use as a local anesthetic. The Peru Indians also found that by drinking a bitter brew made from the bark of the cinchona tree the chills and fevers of malaria could be checked. Cinchona bark contains the drug quinine and, to this day, is the only natural source of it. Indeed, although quinine was synthesized from benzaldehyde, a coal tar product, in 1944, natural quinine is by far cheaper than the synthetic variety.

Morphine. Another well-known drug made by a plant is morphine. Morphine is extracted from seeds of the opium poppy. It is a ***narcotic.*** That is to say it relieves pain and induces sleep, but it has no curative value. Morphine is habit-forming; this is its most harmful effect. Once a person becomes an addict he finds it almost impossible to get along without this debilitating and poisonous drug.

Penicillin, an Antibiotic. Penicillin is a new kind of drug; it is made in the bodies of living organisms. For many years it had been known that "friendly" bacteria in soil destroy virulent bacteria which cause such diseases as typhoid and dysentery. Could friendly bacteria be introduced into the human body to destroy germs that cause such infections as pneumonia and tuberculosis? Such bacteria have been found. ***Antibiotics*** are the chemicals produced by living things which destroy disease-producing organisms. Antibiotics do not, as a rule, kill harmful bacteria. Instead they stifle their life process, thus inhibiting their ability to reproduce. As a result, the body can cure itself.

Penicillin was the first of the antibiotics (it was discovered in 1929) and is still the best known of them. Penicillin is extracted from a green mold (***Penicillium notatum***) that occurs abundantly in soil. It is a crystalline substance which can be safely injected in the blood stream. Penicillin has been used to treat pneumonia, meningitis and many other infections.

Another antibiotic, also isolated from a mold found in the soil, is streptomycin, discovered in 1944. It too is a crystalline solid and has proved effective against many bacteria resistant to penicillin. The antibiotics (which also include aureomycin, terramycin, and chloromycetin) have been so effective against skin abscesses, ulcers, and infectious virulent diseases that they are often called "wonder drugs."

Most antibiotics come from molds found in the soil. Chemists have taken their molecules apart to understand their structures. After 15 years of research penicillin was finally synthesized in 1957. This is a significant step. Chemists can now create many new varieties of penicillin not produced by the mold, some of which may be even more effective against infectious diseases than natural penicillin.

Sulfa Drugs. Sulfa drugs do not occur in nature. They are entirely synthetic, derived from benzene, a coal tar product. The first, sulfanilamide, was made in 1936. Since then dozens of sulfa drugs have been synthesized but the best known and most effective are sulfanilamide, sulfapyridine and sulfathiazole. The sulfa drugs are used to combat such dreaded infections as pneumonia, meningitis, dysentery, typhus, and typhoid.

Laxatives. Laxatives are drugs which stimulate the wave motion (peristalsis) of the intestine and thus relieve constipation. Phenolphthalein, the chemical indicator, is a widely used ingredient of laxative preparations. Emulsions, pills, and even chocolate candy serve as vehicles for phenolphthalein.

Magnesium sulfate (Epsom salts) as well as several other salts are more pronounced in their peristaltic action than phenolphthalein. Such substances are called purgatives. Castor oil, obtained from the castor bean, is one of the best-known purgatives.

Acetylsalicylic Acid, the Headache Pill. This compound is one of the most popular and least harmful pain-relievers (or analgesics in medical parlance). Although it is a pain-killer, a fever-reducer, and an anti-rheumatic drug, little is known of the way it acts in the body. It is by far the most widely used drug. In a recent year no fewer than 12 billion acetylsalicylic acid tablets were consumed in the United States. This is the equivalent of 6000 tons. While this is a relatively safe drug, it should be borne in mind that the indiscriminate use of it may mask symptoms of serious diseases.

STIMULANTS AND TRANQUILIZERS

New drugs which affect the emotions are now being widely used. Some are used to invigorate; they are ***stimulants.*** Others are used to tone down; they are ***tranquilizers.***

Stimulants are usually made from amphetamine and related compounds. Amphetamine is used as a nasal spray to give relief from colds and hay fever. When used in tablet form and sold under such trade names as Benzedrine and Dexedrine, it is a potent stimulant. Apparently it dulls the sense of fatigue and gives patients a feeling of well-being. It also stimulates physical activity, and has been used by athletes for this purpose. Such drugs should never be used unless prescribed by a physician. They have been known to disturb one's personality.

One of the oldest tranquilizers is tobacco. In medicine tranquilizers are called ataractics and the new drugs are sold under such trade names as Serpasil, Equanil and Thorazine. Tranquilizers are administered to relieve patients suffering from mental stress. Although, in many cases, they reduce anxiety they do nothing to affect the underlying psychosis which causes the anxiety.

ANTISEPTICS

Sulfa drugs and antibiotics can destroy harmful germs ***inside*** the body without permanently injuring the body itself. To prevent the infection of open wounds by bacteria, suitable chemicals are often used. Such chemicals are called ***antiseptics.***

There are dozens of antiseptics. Probably the most familiar antiseptic is tincture of iodine, a solution of iodine and potassium iodide in alcohol and distilled water. (The potassium iodide has no antiseptic properties, but it makes the iodine more soluble.) When applied to a wound, the alcohol (an antiseptic) evaporates, leaving behind finely divided iodine, which is also capable of destroying bacteria. Iodine is exceedingly poisonous; it should never be taken internally.

Boric acid (H_3BO_3) is mild and non-poisonous. It may be used as an eyewash.

Hydrogen peroxide (H_2O_2) is really a solution of 3 per cent hydrogen peroxide and 97 per cent water. A trace of blood causes hydrogen peroxide to decompose and froth. The bubbles are oxygen which destroy bacteria and carry foreign matter to the surface of the wound,

$$2H_2O_2 \rightarrow 2H_2O + O_2$$

Carrel-Dakin solution (also called Zonitc) is a solution containing sodium hypochlorite (NaClO). When this solution decomposes it releases chlorine and oxygen, both of which attack bacteria. This solution is particularly good for irrigating deep wounds.

Other familiar antiseptics are mercurochrome, argyrol, chloramine, and metaphen. We should not forget, however, that the safest antiseptic is soap and water. It should be used to remove dirt and grease from a wound before any other antiseptic is applied.

DETERGENTS

Soap is a cleansing agent. How does it act as a cleanser? You will recall that soap is the sodium salt of a fatty acid; its chemical name is sodium stearate. The sodium end of its molecule has an affinity for water but not for oil; the fatty acid radical, however, has an affinity for oil and not water. As a result, soap molecules arrange themselves so that the radical end is in oil and the sodium end in the surrounding water, as shown in Fig. 23-6. In this way a soap barrier is formed which prevents oil droplets from making contact with each other. Instead, oil droplets are dispersed throughout the water; that is, an ***emulsion*** is formed. The emulsifying action of soap can be shown as follows.

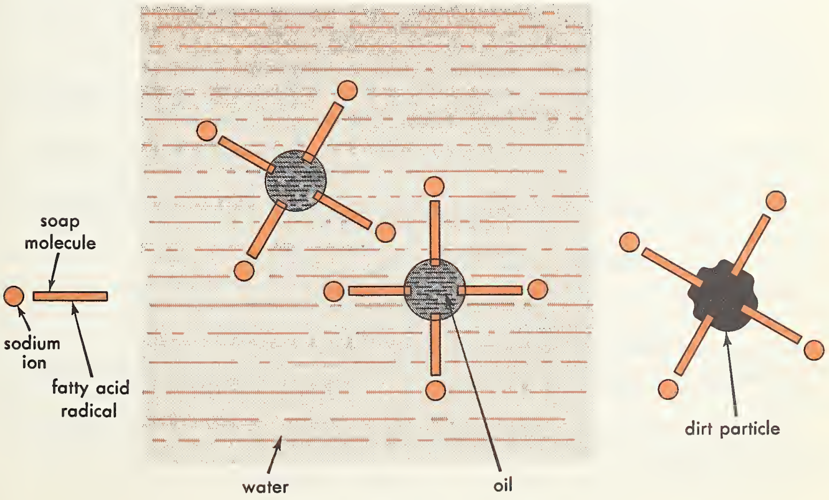
Demonstration

23-1. The emulsifying action of soap

Into each of two test tubes pour about 20 ml. of water and 3 ml. of kerosene. Add some soap solution to one of the tubes. Shake both tubes and let the mixtures settle. Notice that the kerosene and water separate at once into two layers, whereas a more permanent emulsion is formed in the tube containing soap.

It is this emulsifying action of soap that accounts for its cleansing property. As you know, water runs off a greasy surface without wetting it. But with soap, water penetrates the fibers so that the grease is emulsified and carried away.

Fig. 23-6. Soap is adsorbed by grease and dirt, which accounts for the cleansing action of soap.



Dirt particles adsorb soap in the same way as oil droplets do (see Fig. 23-6). As a result, dirt particles are pushed away from the skin and can be rinsed off by water.

Soap is a detergent. The word ***detergent*** simply means cleansing agent. The term, however, is usually applied to the numerous synthetic cleansers that have recently appeared under various trade names.

What are detergents? Detergents are quite different from soaps in chemical composition; they are not made from animal fats. One of the commonest is formed by combining long-chain alcohols and sulfuric acid. Its chemical name is sodium lauryl sulfonate, and the basic material for this detergent is coconut oil.

Are detergents superior to soap in cleansing action? Like soaps, detergents are wetting agents and they emulsify fats and oils. Their great advantage is that they lather well in hard water, even in sea water.

You will recall that calcium and magnesium salts in hard water form a scum with soap (see page 165), which sticks to fabric and is very objectionable. Detergents do not form precipitates with hard water; their cleansing power is just as effective in hard water as in soft. This difference between soap and detergents is shown in the following experiment.

Demonstration

23-2. *The effect of hard water on soap and detergent*

Make some hard water by dissolving about 5 g. of calcium chloride in about 200 cc. of water. Pour some of the solution in each of two tall test tubes. Add a little soap solution to one and liquid detergent to the other. Notice the heavy precipitate in the tube containing the soap solution. Shake the tubes and notice that the detergent gives lots of suds, whereas the soap gives very little.

Things to Remember

A ***balanced diet*** includes carbohydrates, fats, proteins, minerals and vitamins.

The most common ***carbohydrates*** are sugar and starch.

Proteins contain nitrogen as well as carbon, oxygen, and hydrogen.

All green vegetables contain ***minerals*** which are needed for various body functions.

Vitamins are body regulators. Vitamin A was the first vitamin to be isolated—the date, 1914.

Food can be preserved by canning, refrigeration, the addition of preservatives, and irradiation.

Milk contains all the essentials for a complete diet.

Antiseptics prevent the infection of open wounds.

A ***detergent*** is a cleansing agent.

Bacteria and yeasts are microscopic plants.

A ***narcotic*** is a drug that relieves pain and induces sleep, but has no curative value.

Antibiotics are chemicals produced by living things which destroy disease-producing organisms.

Penicillin was the first antibiotic (discovered in 1929).

Questions

GROUP A

1. What is a balanced diet? What does it include?
2. Explain the terms (*a*) sunshine vitamin, (*b*) thiamin, (*c*) carotene.
3. (*a*) In what important respect do bacteria differ from ordinary plants? (*b*) What is the outcome of this difference?
4. Name three common methods of preserving food.
5. What is meant by (*a*) pasteurization of milk, (*b*) irradiation of food?
6. What is a narcotic? Give an example.
7. Some drugs are called ***tranquilizers*** and others ***stimulants***. What are they?
8. Explain the terms ***antibiotic*** and ***antiseptic***. Give two examples of each.

GROUP B

9. What are (*a*) carbohydrates, (*b*) proteins? Discuss their importance as foods.
10. Write briefly on the value of minerals and vitamins in the diet.
11. What is the source and purpose of the following: vitamins A, B, C, and D?
12. Explain the canning process. What steps are taken to exclude and destroy bacteria?
13. What is penicillin? How is it obtained? How does it operate as an antibiotic?
14. What is a detergent? How do modern detergents differ from soap?

Things to Do

Digestion of starch · During the digestive process, large starch molecules are broken down to small glucose molecules, small enough to pass through the intestine walls and into the blood stream. The decomposition of starch can be shown in a simple way.

Put two or three lumps of starch in a 400 cc. beaker of water and stir it until the solution is opalescent. Half fill a test tube with starch solution and add a drop or two or tincture of iodine. A starch-iodine compound is formed which is deep blue in color. This blue color with iodine is used as a test for starch.

Now pour about 150 cc. of starch solution in each of two beakers. Pour about 5 cc. of dilute hydrochloric acid into one and treat the other as follows. Collect a spoonful of saliva by spitting into a spoon and stir it with the starch solution.

Pour about 10 cc. of each starch solution into a test tube and add one drop of tincture of iodine. Repeat this test every minute or two. In the saliva solution the blue color becomes paler and disappears altogether in a few minutes. No color change is observed in the acid solution. Now boil the acid solution and continue testing with iodine. Eventually the blue color disappears.

What has happened? Saliva contains small quantities of a protein called ***ptyalin*** which accelerates the decomposition of starch with astonishing rapidity. Ptyalin is an enzyme, an organic catalyst.

Starch can also be decomposed by mineral acids. But even hot acids act upon starch much more slowly than ptyalin.

Electricity

LOOKING AHEAD

Merely by turning a switch we can operate lighting, heating or power-driven electrical devices. All these devices depend upon the ability of electricity to flow through a wire. This is its most important property. Current electricity is a relatively recent discovery. However, static electricity and the related phenomenon magnetism were known to the ancients more than 2000 years ago.

Modern electricity dates from 1820. This was the year that Hans Oersted discovered that an electric current produces a magnetic field. This important discovery led to the invention of the electromagnet, a device that depends upon the magnetic effects of a current.

The conversion of magnetism to electricity, the reverse of the Oersted effect, was a still more momentous discovery. This principle of electromagnetic induction was discovered independently by Henry (American) and Faraday (English) in 1824, and out of it has come some of our most useful inventions. One is the transformer, which makes the distribution of electric power possible. Without transformers it would be economically impracticable to send large amounts of electric power through wires more than a mile long.

But the biggest impact of all came when electric power was used as the servant of man; that is to say, when the motor was invented. Comparing today's standard of living with that of ancient Greece it has been said that the electric motor has put at the disposal of the average household the power equivalent of 50 or 60 slaves. But motors must be fed with electricity, and before they could operate on a large scale an economic source of electricity was necessary. This was supplied by a generator.

Generators and motors work on the same principle, but produce opposite effects. Generators transform mechanical power into electrical power; motors transform electrical power into mechanical power. Both are necessary in solving our power problem. They provide a convenient means of distributing the power provided by nature. The power of falling water or of a steam engine can be put to many useful purposes via the generator and motor. Together they have made a notable contribution to our way of life by stimulating the demands for motor-driven machines.

← *Masts support high-tension lines which carry electricity at a high voltage from a generator station to a town or factory, often miles away.* (Ontario Hydro)

24

MAGNETISM

MAGNETISM has been known for at least 2000 years. It is not surprising, therefore, that there is uncertainty about the origin of the word **magnet.** One ancient writer says it arose because a shepherd named Magnes investigated the mysterious pull between the iron tip of his crook and a certain kind of rock (an oxide of iron) which we now call magnetite. Another says that the word is derived from Magnesia, a locality in Asia Minor where the rock was first discovered.

Modern scientists have learned how to put magnetism to useful purposes. Both generators and motors, for example, operate because of magnetic effects. One of the earliest uses of a magnet was as a compass. It is said that the ancient Chinese used a suspended piece of magnetite as a compass almost 2000 years ago. That is why magnetite was originally called ***lodestone,*** a word meaning ***leading stone.***

Magnets. What is a magnet? If you rub one end of a steel needle on a lodestone, the needle becomes magnetized. Try this, and then test both the lodestone and your magnet with small iron nails or iron filings. One definition of a magnet is that it is a piece of iron or steel that will pick up iron or steel objects or, we should say, magnetic materials. Today man has been able to make stronger magnets than those made of steel alone. Alnico, for example, is a magnet composed of iron, aluminum, nickel and cobalt. It will lift a piece of iron or steel more than 60 times its own weight.

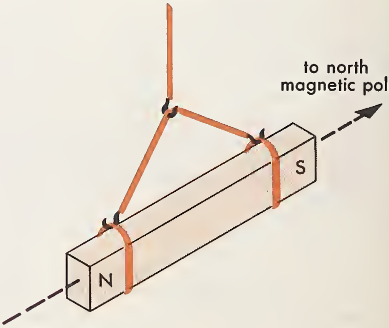
Several demonstrations will show how magnets behave, and we can learn a number of important scientific principles from observing them.

Demonstration

24-1. The direction in which a freely suspended magnet points

Suspend a bar magnet so that it is free to turn, as in Figure 24-1. Observe the direction in which the *N-pole* points; and also the *S-pole.*

Fig. 24-1. A freely suspended bar magnet points in a north-south direction.



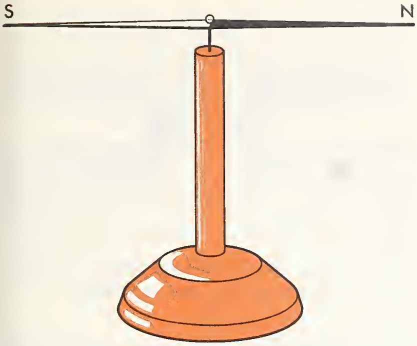


Fig. 24-2. A compass is a magnetized needle supported on a pin point.

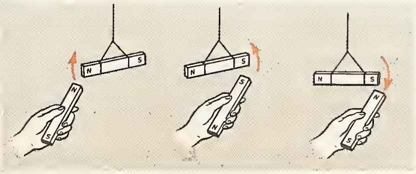
A ***freely suspended bar magnet points*** in a ***north-south direction.*** (The ***N-pole*** means a north-seeking pole, shortened to north pole, and the ***S-pole,*** a south-seeking pole or south pole.) A magnet used to indicate north and south directions is called a compass. Magnetic poles affect each other as will be seen in the following demonstration.

Demonstration

24-2. Magnetic attraction and repulsion

Bring the N-pole of a bar magnet near the N-pole of one freely suspended. (See Fig. 24-3.) Observe what happens. Notice that the suspended magnetic pole moves away from the other magnetic pole. Repeat with the S-pole of the two magnets. Observe. They, too, move away from each other. Now bring the N-pole of the fixed magnet near the S-pole of the suspended magnet. Note that the suspended pole moves toward the fixed pole. Summarize your observations.

Fig. 24-3. Like poles repel; unlike poles attract.



Stated simply: ***Like magnetic poles repel, unlike magnetic poles attract.*** This statement, called the law of magnetism, plays an important role in the theory of the operation of the electric motor, as we shall see later.

Demonstration

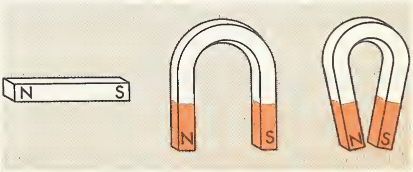
24-3. Do magnets attract all things?

Bring a bar magnet near such objects as a nail, steel pen point, paper clip, bits of copper, brass, aluminum, cobalt, nickel, rubber, paper, zinc, glass and wood. Which are attracted to the magnet, or rather, which materials are magnetic?

So far we have been discussing magnets that are bar-shaped. We have found out that certain substances are magnetic, and that magnetic materials are used to make magnets. Magnets may be made in many sizes and in various shapes such as a U, V, horseshoe, bar and ring. The different shapes are merely for convenience in using in various devices. If the poles are bent toward each other, as in the horseshoe magnet, both poles can apply their lifting force to the same object, hence the lifting force increases.

Magnetic Fields. What is meant by the magnetic field of a magnet? To investigate this problem we can perform an interesting demonstration.

Fig. 24-4. Magnets may be made in various convenient shapes.



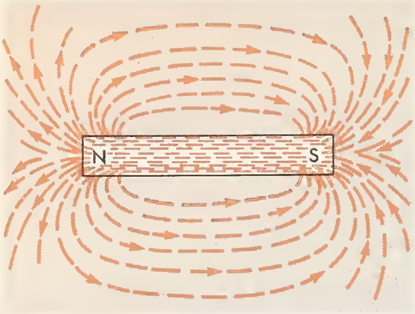


Fig. 24-5. A magnetic field (shown by lines of force) surrounds a bar magnet.

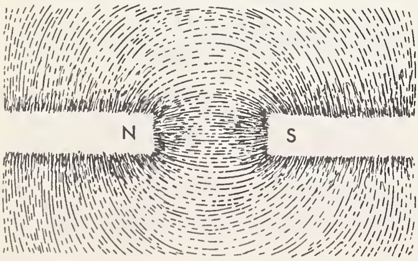
Demonstration

24-4. Magnetic fields

Sprinkle iron filings on a glass plate. Place this plate over a bar magnet. Tap the plate gently. Observe what happens. Repeat this experiment with two bar magnets whose poles are one or two inches apart, placed first with their N and S poles facing each other and then with their N poles opposite each other. Observe the direction of the lines of force of the magnetic fields.

You will notice that when the plate is tapped, the iron filings line up in an orderly manner between the north and south poles. These long curved lines between the poles indicate the *magnetic lines of force of the magnetic field*. The *magnetic field* is defined as the space around a magnet in which it has the power to attract magnetic materials. The magnetic field completely surrounds the magnet and includes *all* the *magnetic lines of force*. (See Figure 24-5.)

Fig. 24-6. Lines of force pass from the N-pole of one magnet to the S-pole of an adjacent magnet.



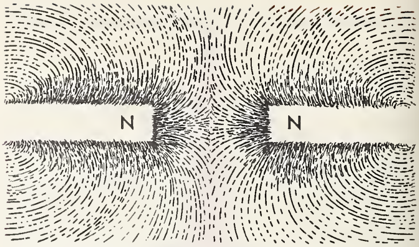


Fig. 24-7. Magnetic lines of force between two poles of the same kind repel each other.

The results of the second part of the demonstration are shown in Figures 24-6 and 24-7. Note that when the north pole of one bar magnet is placed in a straight line with and one or two inches from the south pole of a second bar magnet, the lines of force are again indicated by the orderly line-up of the iron filings between the two poles. However, the lines of force repel each other when the magnetic poles are of the same kind. Here the lines of force bend back to the opposite poles of their same magnets. A check with a small compass placed at various positions along a line of force will indicate that the *lines always go from a north pole to a south pole.*

The Earth Is a Magnet. The earth, a huge spherical mass, behaves as though it were a large magnet. It has one magnetic pole in the Arctic region and another in the Antarctic region. Testing with a compass, we find the north pole pointing to the Arctic region. Hence, by applying the law of magnetism, the magnetic pole in this area is a south pole and the magnetic pole in the Antarctic region a north pole. (See Figure 24-8.) These magnetic poles are more than a thousand miles away from the geographic poles of the earth. Therefore, in using a compass for navigation, corrections have to be made. Charts showing the declination of the compass are furnished to mariners. ***Declination is the angle between the geographic north-south direction and the direction indicated by a magnetic compass.*** (See Fig. 24-9.)

We do not know exactly what causes

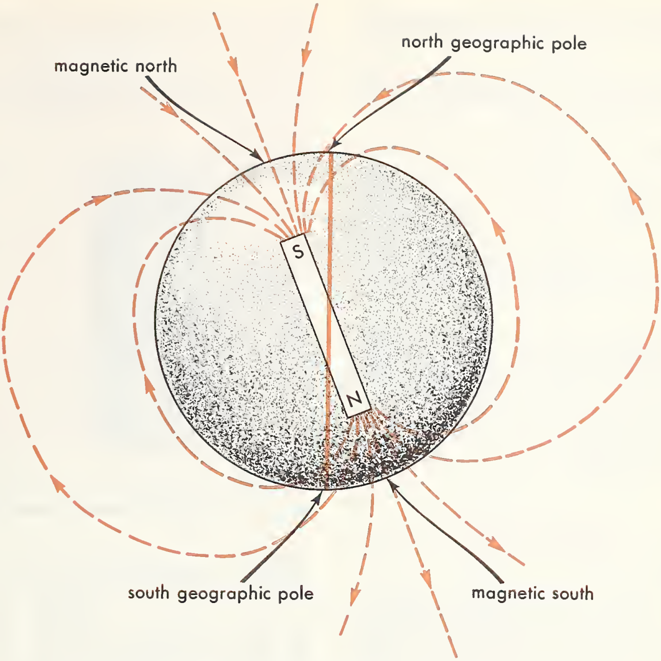
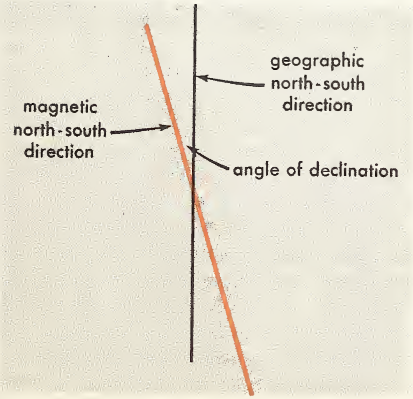


Fig. 24-8. The earth behaves as though it were a huge magnet.

the earth's magnetism. Scientists have been studying the problem for many years, but the answer is not yet known. Perhaps at some time in the future we will know the whole story.

Fig. 24-9. Declination is the angle which a compass needle makes with the geographic meridian.


geographic north-south direction
magnetic north-south direction
angle of declination

A soft iron rod, held in a north-south direction in the earth's magnetic field, can be magnetized by tapping it with a hammer. Try this and test the ends for polarity with a compass. How is it possible to magnetize the rod? This can be explained by the domain theory of magnetism.

Magnetic Domain Theory. Many years ago, Pierre Weiss, a French scientist, said that iron is composed of many small magnetized regions, called "domains." Normally these domains point in every possible direction, so that they cancel one another and the piece of iron is not magnetized. The domains themselves consist of many millions of atoms. The iron becomes magnetized when an outside force lines up the domains in the same direction.

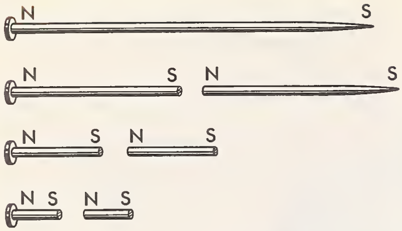


Fig. 24-10. If a magnetized knitting needle is broken, poles appear at the breaks.

If the domain theory of magnetism is correct, small parts of a magnet, consisting of many domains, should also be magnetized and would exhibit polarity. A simple demonstration will offer proof of this statement.

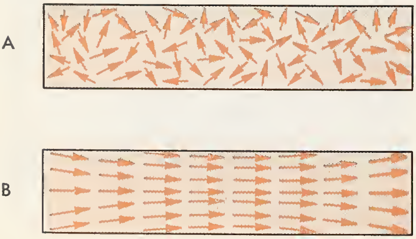
Demonstration

24-5. Theory of magnetism

Take a magnetized knitting needle and break it into two pieces. Test the ends of each piece with a compass for polarity. Break one piece in half again, and test for polarity. Repeat until you have a very small part of the needle left. Continue to test each piece of the needle.

In this demonstration you found that each piece of the magnet is, in turn, a magnet and has a north and south pole. Regardless of how small the piece is, it still is a complete magnet with two poles. This means that when a ***magnetic material is magnetized,*** as in the case of the knitting needle, the ***domains*** composed of millions of atomic magnets are ***lined up*** so that their ***north*** poles are all pointing in the same direction. This makes the total piece a strong magnet. (See Fig. 24-11.)

Fig. 24-11. If iron is unmagnetized, its atomic magnets are in disarray; if iron is magnetized, its atomic magnets are lined up as in the lower picture.



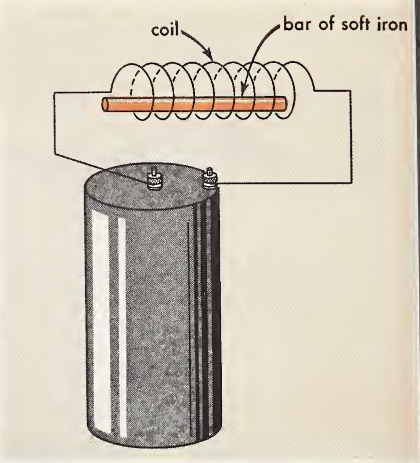


Fig. 24-12. An iron bar can be magnetized by placing it in a coil through which a direct current flows.

How Are Magnets Made? There are several ways to accomplish the orderly line-up of the small-domain magnets in order to make a strong magnet. One way is to stroke the magnetic substance with the poles of a ***permanent*** magnet. Another way is to tap it with a hammer, as mentioned before. A still better way is to place the magnetic material within a coil of wire through which a direct current of electricity is flowing.

If the substance to be magnetized is made of hard steel or of an alloy of iron, aluminum, nickel and cobalt, a strong permanent magnet will result. Once the magnetic domains are lined up it is difficult for them to vibrate out of line, because of the strong force between the molecules of the hard material. Structurally softer materials, like soft iron, will not retain their magnetism very well.

Removing Magnetism. How can a magnet be made to lose its magnetism? To do this it is necessary to break up the orderly domain alignment. This can be done in several ways. By striking a magnet, its domains can be jarred out of line and some of the magnetic strength removed. Then, again, if you place the magnetized material within a coil of wire and send an alternating current of electricity through the coil, the domains in the magnet will change their orderly orientation. A third way is to heat the magnet, thereby increasing the energy of the atoms and causing them to break their formation within the domains and the domain formation within the entire structure of the magnet.

Sometimes it is necessary to remove magnetism from a steel piece, such as the hairspring of a watch. As already stated, placing the watch inside a coil of A. C. electricity will cause the domains to break their alignment, thus demagnetizing the spring.

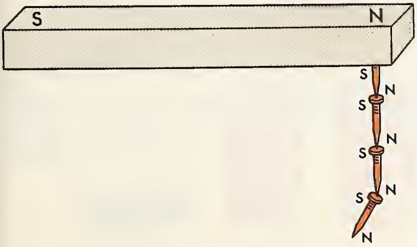
Magnetic Induction. Magnetic induction can be shown by means of an experiment as follows:

Demonstration

24-6. *Magnetic induction*

Support a small nail from the pole of a bar magnet. Then support another nail from the end of the first nail. Repeat until you have as long a string of nails as the magnet will hold.

Fig. 24-13. The nails attract each other because of magnetic induction.



How is it possible to hold the last nail? The first nail is a magnet, since the bar magnet causes the domains of the nail to line up according to the domain theory of magnetism. In turn each nail magnetizes the one below it by means of its own magnetic field. The magnetism induced is only temporary, lasting as long as the first nail is within the magnetic field of the bar magnet. Removed from this magnetic field, the nails soon lose their induced magnetism. The ***process of temporarily magnetizing a substance in a magnetic field is called magnetic induction.***

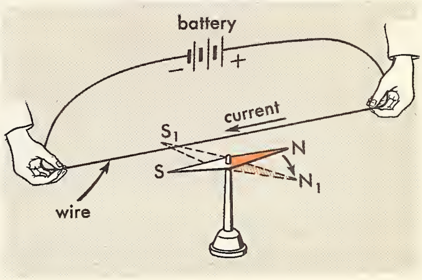
The Magnetic Field of a Current. Hans Christian Oersted, a professor of physics at the University of Copenhagen, Denmark, made a very important discovery in 1819. While lecturing to a science class he inadvertently placed a wire, carrying an electric current, parallel to a compass needle. To his surprise, the needle turned to a new position nearly perpendicular to the wire. When Oersted reversed the direction of the current, the needle also reversed to a similar position. See Fig. 24-14.

Demonstration

24-7. *Oersted's experiment*

Connect six feet of insulated wire through a knife switch to three dry cells connected in series (positive terminal of one to negative terminal of next, etc.). Place a compass near the wire and hold the wire parallel to

Fig. 24-14. Oersted's experiment. A wire carrying a current shows a magnetic effect.



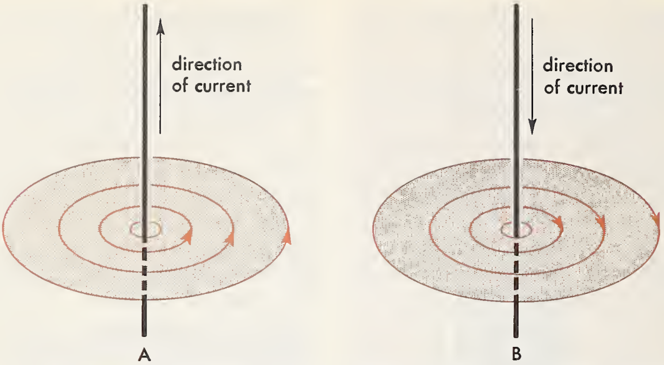


Fig. 24-15. The magnetic field about a current is a series of concentric circles around the wire.

and above the compass needle. Close the switch and observe what happens. Repeat, holding the wire under the compass needle. Reverse the direction of the current flow and repeat. Summarize your results.

This experiment shows that there is a ***magnetic field surrounding a wire carrying an electric current.*** This ***magnetic field is a series of concentric circles about the wire*** as shown in Fig. 24-15.

An Electromagnet. So far we have been concerned only with a permanent magnet, a substance that is a magnet all the time. But a permanent magnet is not as useful as one that could be magnetized and demagnetized at will. Let us consider this type. An electromagnet consists of a soft-iron core and a coil of insulated wire. You can make an electromagnet and do an interesting experiment with it.

Demonstration

24-8. Making an electromagnet

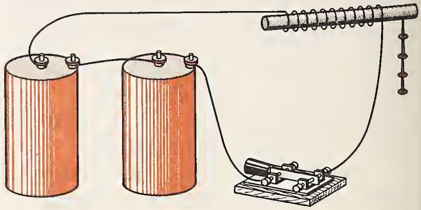
Take a large iron bolt, some insulated copper wire, a knife switch, and several dry cells. You will also need some small iron nails. Wind about 50 turns of wire around the bolt, and connect the ends of the wire through the switch to a dry cell, as shown in Fig. 24-16.

Throw the switch and bring an end of the bolt near some small iron nails. Observe what happens. How many nails will your electromagnet pick up? Count them. Turn off the switch. What happens? When does the iron core of the electromagnet become a magnet? Check the polarity of each end of the core with a compass.

Repeat the above experiment with more turns of wire around the bolt. Count the number of nails held by the electromagnet. Repeat with several dry cells, connected in series with positive terminal of one attached to the negative terminal of the next, etc. What are your results in both cases?

You will note that, whenever the current flows, the bolt is magnetized by the magnetic induction of the magnetic field of the coil; whenever the current is shut off the bolt quickly loses its magnetism. A temporary magnet results which can be easily controlled. Note that the electromagnet can be made stronger by (1) increasing the number of turns of wire in the coil, and (2) by increasing the

Fig. 24-16. Making an electromagnet.



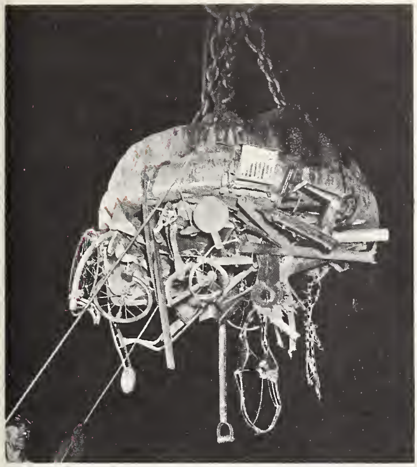


Fig. 24-17. Scrap iron is being transferred on an electromagnet by means of a crane to a steel furnace.

electric current used. A softer iron core would also help to increase the magnetic strength, since its domains would line up more readily when the current is on and vibrate out of line more readily when the current is turned off.

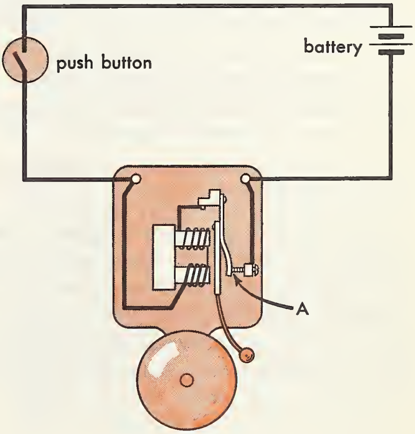
Electromagnets are used in many ways,—in electric motors, generators, in the telegraph, in bells, television sets, radios, buzzers, telephones, and electric door chimes. Large electromagnets are used for transporting heavy iron objects from one place to another as shown in Fig. 24-17. Here a huge electromagnet is attached to a crane. A heavy current through the coil produces the strong magnetic field needed to lift heavy objects. Sheet steel, rails, and scrap iron are moved in this manner. When the current is turned off the coil is demagnetized, and the objects fall.

An Electric Bell. The door bell or buzzer is a familiar example of an electromagnet. A circuit for a door bell is shown here in Fig. 24-18.

When the button is pressed, the electric circuit is completed and current flows through the coils of the bell, making their cores a magnet. This magnetism pulls the armature over toward the core causing the attached clapper to hit against the gong. The circuit is broken when the armature spring breaks contact at point *A*, the current stops flowing and the electromagnet is de-energized. The armature is released to move back through the spring action of the strip of spring steel that supports it. Contact is again made and the electric circuit is completed once more. The whole action is repeated, again and again, causing the bell to ring.

If there is no clapper, a buzzing sound is produced instead of the ring. In an electric chime, the button is pressed, a hollow coil is energized, drawing a suspended iron plunger into its end. This plunger flies through the coil, striking the gong to produce the tone. When the circuit is broken, the current stops and the plunger returns to its original position.

Fig. 24-18. Can you explain why the bell rings when the button is pressed?



Things to Remember

Like magnetic poles ***repel*** each other; ***unlike*** poles ***attract.***

A ***magnetic field*** surrounds a magnet.

A ***magnetic compass*** is a small bar magnet that is free to rotate.

The north pole of a freely suspended magnet points to the ***earth's magnetic south pole*** (in the Arctic).

Compass readings need to be corrected for ***declination.***

A ***magnetic field*** surrounds a wire carrying an electric current.

An ***electromagnet*** consists of a coil of wire (insulated) wound around a soft iron core.

The ***strength of an electromagnet*** can be increased by increasing the number of turns in the coil and by increasing the current used.

Questions

GROUP A

1. State the law of magnetism.
2. What proof do you have that the earth acts like a magnet?
3. How would you make a strong permanent magnet?
4. What is the advantage of bending a magnet in the shape of a horseshoe?
5. Describe an experiment that shows that like poles repel and unlike poles attract.

GROUP B

6. Show that the earth has a south magnetic pole in the Arctic region.
7. Draw a sketch of the magnetic field about a bar magnet. Label the poles.
8. Tell how you can plot the magnetic field.
9. Name five different devices that use electromagnets.
10. Define the term magnetic declination and discuss its importance.

Things to Do

To make a strong electromagnet · Take a soft-metal rod (iron), or preferably, a large iron bolt. Make two cardboard washers and fit them over the bolt, one at the head end and the other at the nut. Now wind the bolt with as many layers of magnet or bell wire (insulated) as you can accommodate to the top of the washers. Wind neatly and, after completion, cover the top of wire with tape to hold the wire in place. Using two or three dry cells, test the strength of your electromagnet.

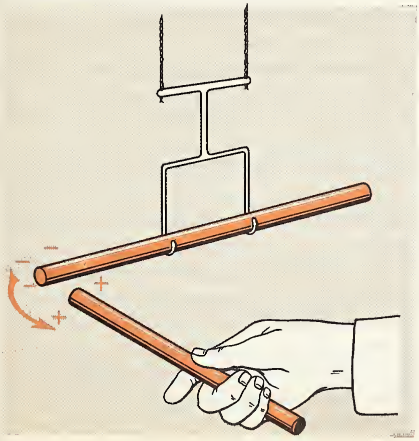
25

STATIC ELECTRICITY

You have probably experienced static electricity in petting a cat, combing your hair, or shuffling your feet across a rug on a dry day. Try this latter, keeping your feet in contact with the rug. Then touch a metal doorknob or touch your finger tip to the ear lobe or nose of a friend. Notice the spark. He will feel the sharp prick of the ***static electricity*** (electricity stored, or at rest) jumping from your finger. When you rubbed your feet across the rug your body picked up negative charges from the rug and, by contact, you transferred some of these charges to your friend.

Electricity from Friction. The ancient Greeks found that the substance amber, when rubbed with cloth, attracted small bits of hair and other materials. You can do a similar experiment. Take a polystyrene or Bakelite rod. Charge it by rubbing it with a piece of fur or wool. Bring the rod close to some torn bits of paper. Observe what happens. At first the paper is attracted and then it is repelled. Take a glass or Lucite rod and charge it by rubbing with a silk cloth. Repeat as above with the small bits of paper. It behaves like the rubbed Bakelite rod. A simple demonstration will bring out the difference between the charges on the two rods.

Fig. 25-1. A negatively charged rod is attracted by a positively charged rod.



Demonstration

25-1. Difference between charges

Suspend a charged polystyrene rod in a stirrup as in Fig. 25-1. Bring the charged end of another charged polystyrene rod near the charged end of the suspended rod. Note that the freely moving rod is *repelled.* Charge a Lucite rod and repeat the demonstration. Note that the polystyrene rod (suspended) is *attracted* to the Lucite rod. Why? Why was the suspended rod attracted in one case and repelled in the other? (*Note:* This experiment works best in dry air—moisture causes the charges to leak off.)

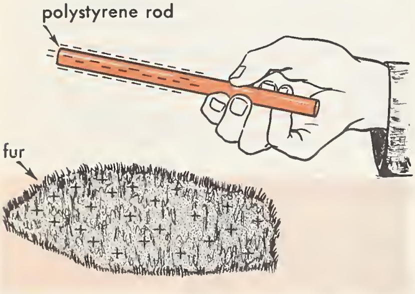
To understand how the rods are charged in the rubbing process, and why they attracted or repelled each other, it is necessary to understand something of the structure of matter.

Electrons and Protons. All matter is made up of invisible small particles called ***atoms.*** The ***normal atom*** contains negatively charged ***electrons*** and positively charged ***protons*** in equal numbers and hence is electrically neutral, since the negative charge of a single electron is equal to the positive charge of a single proton. The mass of the electron is small; that of the proton about 1800 times larger, so that electrons move about more readily than protons.

How does the polystyrene rod get its negative charge? In charging the polystyrene rod by rubbing it with fur, some electrons were transferred from the fur to the rod, leaving the fur charged positively and the polystyrene rod charged negatively. (See Fig. 25-2.) In other words the polystyrene rod has an excess of electrons. The Lucite rod when rubbed with silk transferred some electrons to the silk, making it negatively charged and leaving the Lucite rod with a deficiency of electrons or charged positively.

Since an electron is so much lighter than a proton, it is easier for electrons to move. You will recall that in an earlier chapter we discussed valence and the atomic structure of the atom, including orbits. In charging a body only the valence electrons (in the outer orbit) are removed, and the protons and all the other electrons are held to neighboring atoms by strong electrostatic forces. The ***charge on a rod is always due to the transfer of electrons. A gain of electrons produces a negative charge; a loss of electrons a positive charge.***

Fig. 25-2. The polystyrene rod picks up electrons from the fur when the two are rubbed together.



We noted in the demonstration that the negatively charged polystyrene rod was attracted to the positively charged Lucite rod, and hence we must conclude that ***unlike charges*** attract each other. The demonstration also proved that rods of like charge repelled each other. The ***law of charges*** can be stated: ***Like charges repel; unlike charges attract.*** This is like the law of magnetism, where like magnetic poles repel and unlike magnetic poles attract. It should be realized, however, that magnetic forces and electric forces are not the same. ***An electric charge exerts no force on a magnetic pole.***

Protection against Static Electricity. There are many examples of electricity produced by friction. The motion of air past the surface of a rapidly moving vehicle charges the vehicle, and in dry weather the rubber tires prevent the charge from leaking off into the ground. A gasoline truck needs to be protected against this static charge. You may have noticed a chain hanging from the truck and dragging on the ground. This chain conducts the charges to the ground.

If you have been driving in an automobile and you stop to buy a paper, as you hand the coin to the newsboy he may touch the metal body of the car and receive a shock. Cars rolling over a toll bridge pass over a flexible wire, which removes the static charge from the car before the toll collector receives the fee. Paper passing through the rolls of a printing press, as in a newspaper plant, builds up a tremendous static charge. This charge is dissipated by fine flexible wires or by a fine spray of water as the paper comes off the press. Machines with moving belts are grounded to remove their static charge. Can you think of other ways in which static electricity is hazardous?

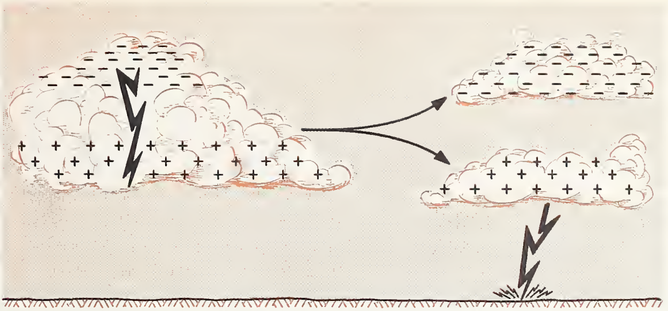


Fig. 25-3. Lightning may jump from oppositely charged parts of the same cloud, or from a charged cloud to the earth.

Insulators and Conductors. We have seen that a polystyrene rod and a Lucite rod can be given a charge by means of friction, the polystyrene rod being negatively charged and the Lucite rod positively charged. These charges tend to remain in a fixed position on the surface of the rods. A metallic rod cannot be charged when held in the hand, since the charges run along the metal to your body. However, if a polystyrene or Lucite rod is used to support the metal, then the metal rod can be charged. ***Metals are good electrical conductors;*** they are substances into or out of which electrons (negative charges) can move freely from atom to atom. Copper and aluminum are examples of good electrical conductors, widely used to carry electricity.

Glass, rubber and porcelain are good ***non-conductors*** or ***electrical insulators,*** and hence are used to prevent the escape of electricity from metallic conductors.

Nature's Electricity. Nature puts on a spectacular display of electric sparks and terrifying rolls of thunder. Lightning is nothing more than ***static electricity of tremendous charge.*** Benjamin Franklin proved this in his famous kite experiment when he drew electricity from the sky during a thunderstorm. *Caution: Do not attempt to do this—it is dangerous.*

How do clouds get their charge? The exact mechanism is not known. One theory states that as air currents move past the water drops of the clouds, they charge them, some positively and some negatively. One of the best hypotheses is that large drops of water are split into small drops in violent air currents and become oppositely charged. It is possible, in a large cloud, for the top to carry one kind of charge and the bottom another. By the law of charges the positive and negative charges attract each other so that a discharge takes place. This discharge may be between parts of a cloud, between two clouds, or between a cloud and the earth. (See Fig. 25-3.)

Let us see what happens between a negatively charged cloud and the earth. This cloud repels electrons on objects projecting above the surface of the earth, causing these electrons to run to ground and leaving the object charged positively. Then a lightning discharge takes place, drawing off negative charges from cloud to object, thereby discharging the cloud to ground. The flash may cause fire or other physical damage, and persons nearby may be shocked or even electrocuted. (See Figs. 25-4 and 25-5.)

Lightning rods are used to protect buildings against lightning. A lightning rod consists of a copper rod, with a sharp point, attached to a flat band of copper that runs down the side of a building as shown in Fig. 25-6. Connected to the copper strap is a copper plate that is buried in the ground. Experiments have proved

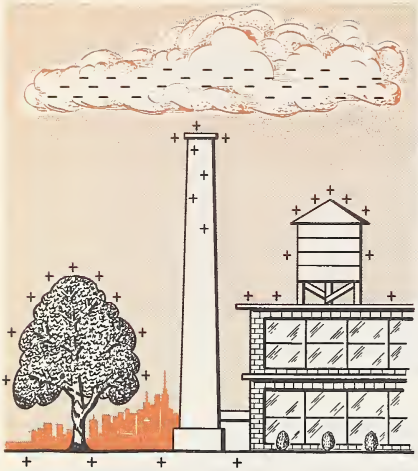


Fig. 25-4. A low-hanging negatively charged cloud induces a positive charge on objects on the ground.

that charges tend to concentrate on points, and that electrons leave points easily. The leakage of electrons from a point can take place in two ways, (1) from ground to cloud or (2) from cloud to ground. If the approaching cloud is charged negatively, the electrons leak from cloud to ground through the lightning rod. If the cloud is charged positively, then electrons will be drawn to the sharp-pointed tip of the lightning rod and leak away into the atmosphere, thus dissipating or neutralizing the positive charge on the cloud. This leakage reduces the chance of a direct flash of lightning, resulting in less danger of damage to life or property.

Fig. 25-5. Two flashes of lightning striking the earth at the same time.

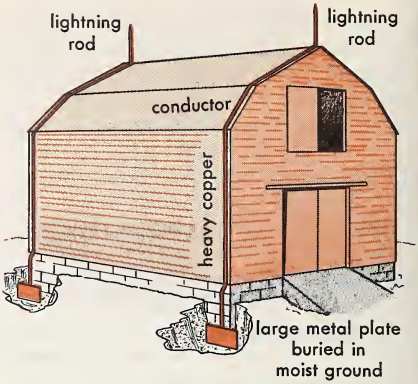


Fig. 25-6. A barn protected by lightning rods.

Power transmission lines usually have a well-grounded single wire running above them to carry off any lightning discharge and thus prevent damage to the equipment or stoppage of electricity to the consumer.

Shelter against Lightning. Perhaps you may wonder where to take shelter during a thunderstorm. Under a tree would certainly be dangerous, especially under a lone tree. Trees are only fair conductors, and the lightning might jump from the tree to your body, taking an easier path to ground. Automobiles, with their steel tops and bodies, are good shields since the electric charges can run over the outside surface of the metal conductor of the car and not through the middle of it. Homes, with their water, gas, and electric lines grounded, are fairly safe shelters. Large buildings are also grounded through their steel frame structure.

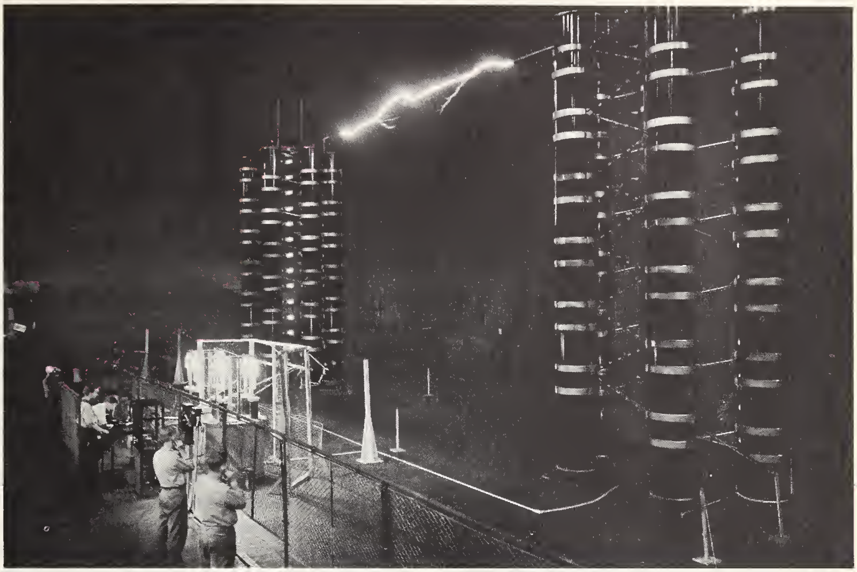
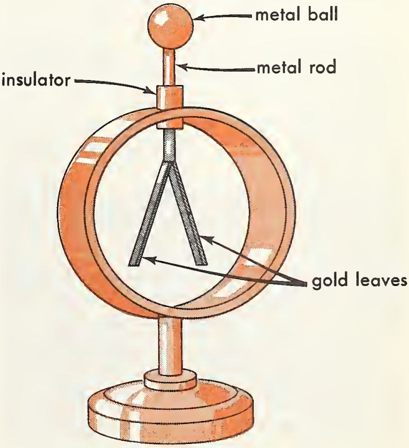


Fig. 25-7. An artificial lightning generator. A current of about 23,000 amperes at a pressure of 10,000,000 volts flows for about a millionth of a second. Notice the men in the foreground are protecting their ear drums against the loud roar.

Man-made Lightning. Man cannot harness the lightning of nature; it is uncontrollable and exceedingly dangerous. However, he seeks to know more about it and to learn how to control it, to prevent damage to electric power lines. Several giant research laboratories have been set up to create artificial lightning and study its effects on various materials. The Van de Graaff generator is an example of a machine used to produce artificial lightning. Some laboratories have been able to obtain voltages of over 15,000,000 volts, causing a charge to leap distances of over 50 feet. (See Fig. 25-7.)

Detecting Electric Charges. An instrument called an ***electroscope*** is used to detect and determine the kind of electric charge on a body. An electroscope is a glass box with an insulated stopper, through which a metal rod is inserted. The top of this rod has a metal ball and the bottom (in the container) is connected to two leaves of thin metal. (See Fig. 25-8.) Aluminum, silver or gold may be used for the leaves of the electroscope,

Fig. 25-8. A gold-leaf electroscope can be used to detect a charge on a body.



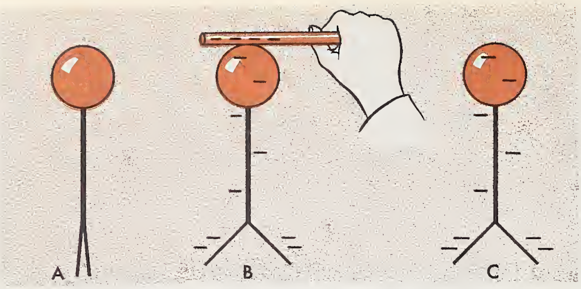


Fig. 25-9. Charging an electroscope negatively by contact.

but gold is usually used since it can be beaten into a very thin foil of little weight. Let us use a demonstration to show how to charge an electroscope by contact.

Demonstration

25-2. Charging an electroscope by contact

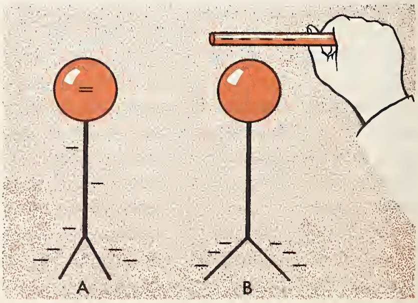
A *neutral* electroscope has the same number of electrons as protons, so that the instrument has no net electrical charge on it. Charge a neutral electroscope by touching a charged polystyrene rod to the ball of the electroscope. Observe. Remove the rod and explain what has happened. Repeat this demonstration, this time using a charged Lucite rod to touch the ball of a neutral electroscope. Remove the Lucite rod and explain what happened.

In the first case, since the polystyrene rod is charged negatively, as we have learned, electrons will flow from the polystyrene rod to the metal ball of the electroscope, which is a good conductor. The ball then has a surplus of electrons which flow down the metal rod to the leaves, causing them to move apart or ***diverge*** since ***like charges repel.*** The ***electroscope will then be charged negatively,*** as shown in Fig. 25-9.

In the second case, the Lucite rod carries a positive charge, as we have learned. Hence, electrons flow from the leaves, via their metal rod, to the ball and on to the surface of the Lucite rod, leaving the leaves with a deficiency of electrons. The Lucite rod has removed ***electrons*** from the electroscope by the ***attraction of unlike charges,*** and when the rod is taken away the electroscope is left charged positively. Since like charges repel, the two positively charged leaves of the electroscope will diverge. Notice that whenever a ***body is charged by contact, the sign of the charge produced is the same as the sign of the charge producing it.***

How is the electroscope used to determine the presence of an electric charge? If an uncharged body is brought near the metal ball of the electroscope, the leaves will not move. However, if a charged body is brought near, the leaves will move apart and if the body is moved away, the leaves of the electroscope will fall together. In this way the presence of an electric charge on a body can be determined. The distance the leaves move

Fig. 25-10. If a negatively charged rod is placed near the ball of a negatively charged electroscope (A), the leaves diverge further as in B.

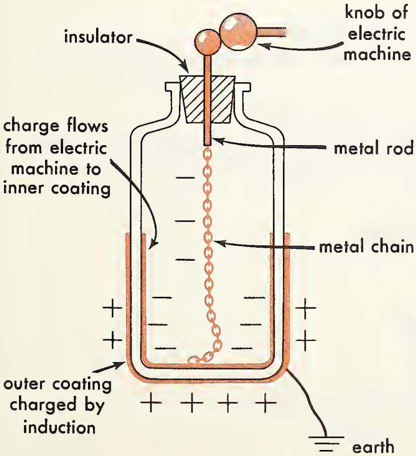


is some indication of the strength of the charge.

If a charged body of unknown charge is brought near the metal ball of a negatively charged electroscope, the leaves will move either farther apart or toward each other. If the leaves of the electroscope move farther apart, we know that the body was negatively charged. How do we know this? A negatively charged body would repel electrons from the ball of the electroscope down to its leaves, making them more negative, hence moving farther apart. If the leaves of the electroscope should move toward each other, then the body must have had a positive charge upon it. Here, electrons would be attracted from the leaves, up through the rod to the ball, because of the attraction of unlike electric charges. The leaves, having lost some of their negative charges, would tend to move together. We have now seen how an electroscope can be used to test the presence of a charge on a body and how the sign of that charge can be determined.

Condensers. As early as 1745, in Leyden, Holland, men stored electric charges in a device called a ***condenser.*** A condenser is made up of two plates of conducting material separated by an insulating substance. This insulation may be air, glass, mica, or even waxed paper. The Leyden jar, so called for the place of its discovery, consists of a glass jar coated inside and out with thin tinfoil, the two "plates." A metal rod running through an insulating top is connected to the inner foil by a chain resting upon the bottom of the jar. The glass serves as the insulating substance, or ***dielectric,*** between the tinfoil "plates." (See Fig. 25-11.)

Fig. 25-11. Charging a Leyden jar condenser.


insulator
knob of electric machine
charge flows from electric machine to inner coating
metal rod
metal chain
outer coating charged by induction
earth

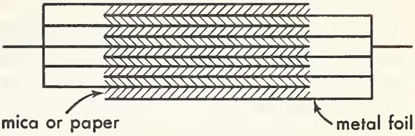
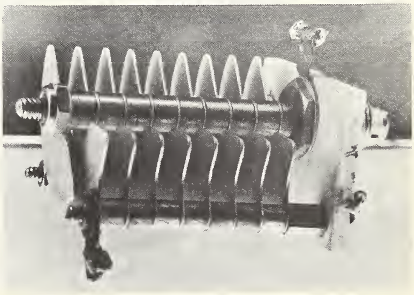


Fig. 25-12. A simple condenser. Strips of metal foil are separated by mica or plastic.

To charge the Leyden jar we first ground the outer foil. Then we bring the negatively charged terminal of a static machine in contact with the metal knob of the jar. In this way negative charges will accumulate on the inside foil, repelling negative charges from the outside foil into the ground. The negative charges on the inside will be strongly held by the positive charges induced on the outer foil coating. A strong charge can be built up. The Leyden jar can be discharged by bringing the inner and outer foils in contact through a pair of discharge tongs with an insulated handle. A sharp spark results.

Fig. 25-13. A variable condenser. By turning the central shaft, the upper plates mesh with the lower ones.



The amount of charge that can be stored in a condenser is called its ***capacity***. The capacity depends on (1) the area of the condenser plates, (2) the distance between the plates, and (3) the nature of the insulator or dielectric between the condenser plates. To obtain a high capacity the area of the plate is ***increased,*** the distance between the plates is ***decreased*** and a dielectric of high value is used. Today, condensers are made of thin foil separated by mica or plastic and stacked or rolled to provide more surface area. (See Fig. 25-12.) Many types of condensers are used in radio, television, telephone circuits, the ignition system of the automobile and in X-ray machines.

Variable condensers enable the operator to change the capacity of the device by turning a movable set of plates, called the ***rotor,*** into a fixed set of plates called the ***stator***. The two sets of plates mesh, but do not touch one another. Air is the dielectric. (See Fig. 25-13.) As the plates are meshed, the capacity of the condenser increases. The tuning knob of your radio set is usually connected to a variable condenser. By changing its capacity, you tune your receiver to different radio stations. The function of the condenser in a radio will be made clear in a later chapter.

Things to Remember

All ***atoms*** contain protons and electrons.

An ***electron*** has a negative charge.

A ***proton*** is about 1800 times heavier than an electron and has a positive charge.

Atoms are usually ***electrically neutral;*** they contain equal numbers of protons and electrons.

A charge on a rod is always due to the ***transfer of electrons;*** a gain of electrons produces a negative charge, and a loss of electrons a positive charge.

Like charges repel each other; ***unlike charges*** attract.

An electrical ***conductor*** is a substance into or out of which electrons can flow freely from atom to atom.

Metals are ***good conductors:*** especially silver, copper, gold, and aluminum.

Electrical ***insulators*** are substances into or out of which electrons cannot move freely from atom to atom.

Glass, rubber, porcelain and plastics are ***good insulators.***

Lightning rods are used to protect buildings from the damage of lightning discharges.

An ***electroscope*** is a device for detecting electric charges.

A ***condenser*** is used to store electric charges and serves a useful purpose in radio, television, X-ray, ignition systems for cars and in telephone circuits.

Questions

GROUP A

1. Explain the operation of a lightning rod.
2. How can you charge a body positively?
3. State the law of charges.
4. What substances are used to conduct electricity? Why?
5. Name some good insulators for electrical circuits.

6. What causes lightning?
7. Why is a condenser used in a radio circuit?
8. Why do gasoline trucks often have a piece of chain dragging on the road?
9. Give several examples of static electricity hazards in industry, and tell how these hazards are reduced.

GROUP B

10. Describe an experiment that proves the law of charges.
11. Describe what happens when a positively charged cloud approaches a lightning rod.
12. Charge an electroscope and use it to detect an unknown charge. Describe what happens and prove the kind of charge of the unknown.
13. Explain the operation of a condenser.

Things to Do

To make a pith ball electroscope · Take a wooden block about 4 inches by 3 inches. In the center, mount a piece of strong wire bent into the shape of a large inverted "J." The wire should be about 9 inches long, and mounted so that the long straight part of the "J" is standing up in the board with the hook part bent down toward the table. Tie a small pith or cork ball that has been dipped in aluminum paint on the end of a piece of silk thread. Bend the end of the wire into a small hook and tie the silk thread to this hook. A tiny piece of cork or a piece of feather down can be substituted for the pith. Charge your electroscope and use it for experimenting with electric charges.

26

CURRENT ELECTRICITY

ELECTRICITY plays a very important role in the lives of all of us. From the time we rise in the morning until we retire at night, the pushing of buttons provides us with light, heat, sound, power, transportation, communication and entertainment. Electricity operates the refrigerator, the vacuum cleaner, the television set, the radio, the telephone, the heater, the home workshop and the door chimes. Moreover, industry uses electrical power to manufacture goods. Planers, lathes, drill presses, milling machines, conveyor systems and many other machines are operated by this friend of man, electricity. An understanding of electricity—its kinds, its production, uses, and safe handling—is advantageous to all consumers.

First of all, where does electricity come from? If we look along the street, we find that wires lead from our homes to a large iron box (a transformer) high up on a pole. If we follow these transmission wires back to their sources, we find that they lead through a substation with another transformer back to a large power plant where an electric current is produced. In the power plant large generators, operated by a turbine powered by the energy of coal or turned by water power, provide us with electrical energy. We shall discuss the complete story of how electrical energy is generated in a power plant in a later chapter.

An electrical current in a wire is a flow of electrons produced by a difference in electrical potential or voltage. Of course, we remember that these electrons are negative charges and that they flow along a conductor. For useful work from electricity, we need to maintain a steady difference of potential and thus to produce a continuous flow of electrons (or current).

Scientists tried for many years to obtain electrical energy in a convenient and practical form. They evolved two devices that are almost universally used today as sources of electricity. These are the generator or dynamo which will be discussed in Chapter 28, and the electric cell which generates electricity by chemical action. Electric cells or combinations of them, called batteries, are used in flashlights, automobiles, airplanes, submarines, and portable radios.

Current electricity is of two types, direct (D. C.) and alternating (A. C.). By direct current we mean the movement of electrons always in one direction. Alternating current is defined as current flowing in alternating directions, rapidly reversing its direction of flow at regular very short intervals. In Canada 60 cycles per second is the most common alternating current. This means that the alternating current changes direction 120 times per second, since a cycle includes the flow first in one direction and then in the reverse direction.

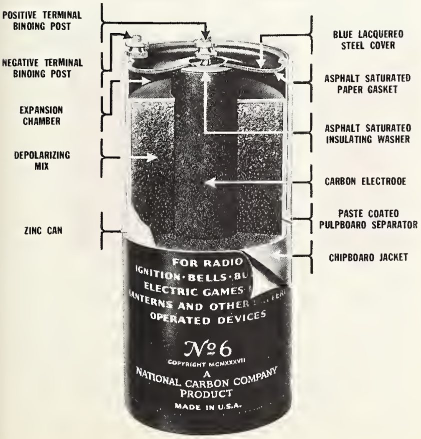
Sources of Direct Current Electricity.

Dry cells and storage batteries are sources of direct current. Later we shall see that electric generators can also supply us with D. C. electricity.

A dry cell consists of a zinc can (the negative terminal) and a carbon center rod (the positive terminal), surrounded by a moist paste of ammonium chloride, granulated carbon and manganese dioxide. The zinc can is lined with blotting paper moistened with ammonium chloride. The top of the cell is sealed with wax, tar, or sealing wax to keep the moisture in the cell. You can see that the dry cell is really not dry but moist. In fact, if it were dry it would not function at all. Terminals are fastened to the zinc and carbon electrodes for ease in connecting wires. A dry cell is shown in Fig. 26-1. If you can obtain an old dry cell, cut away a portion of it and study its structure. Can you identify all of its parts?

Whenever the dry cell is made a part of a complete electrical circuit, a chemical action takes place in the cell. This action develops an electrical force called an electromotive force (electrical potential or voltage) that causes a current of electricity to flow. The electron flow is from the negative to the positive terminal. The current is a steady flow of electrons, lasting until the chemicals of the cell are used up or until the circuit is broken. The normal dry cell will develop a voltage of about 1.5 volts and, when new, will provide a maximum current of 30 to 40 amperes.

Fig. 26-1. A cross section of a dry cell.



It is better for a dry cell to be used intermittently than to be left standing. Action between impurities within the cell and the zinc can cause loss of electrical energy, and the cell deteriorates. The manganese dioxide, in the paste of the cell, oxidizes to water any hydrogen produced in the cell. Hydrogen, a non-conducting gas, increases the resistance of the cell, thereby reducing the current that can be drawn from it.

A Wet Cell. You can perform an interesting experiment producing direct current electricity by means of chemical action in a simple, easily constructed wet cell.

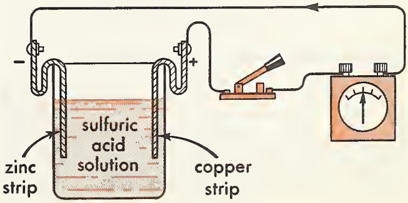
Demonstration

26-1. D. C. electricity by chemical action in a cell

Set up a one-inch wide piece of copper and a like-sized piece of zinc in a glass jar as shown in Fig. 26-2. Fill the jar about two-thirds full with dilute sulphuric acid and connect the wires from the two metal plates to a galvanometer and switch as shown. Close the switch and note what happens to the galvanometer.

When the switch was closed the needle of the galvanometer moved, indicating that an electric current had been pro-

Fig. 26-2. Electricity may be generated by chemical action in a wet cell as shown above.



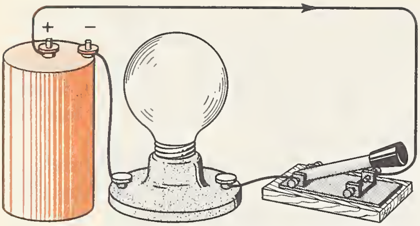


Fig. 26-3. A simple electric circuit.

duced. During the chemical action between the zinc and the acid, electrons from the negative zinc terminal were forced through the galvanometer to the positive copper terminal. Thus a direct current of electricity resulted. The deflection of the needle in one direction indicates D. C. The voltage of this cell is about one volt, but the maximum current flow is small, probably less than one ampere. The current of this simple cell falls off rapidly because of the action of hydrogen bubbles on the copper plate, called polarization. This latter action decreases the voltage of the cell, and hence the current is also decreased.

An Electrical Circuit. What is an electrical circuit? In order to have a flow of electric current there must be a complete circuit. A complete electrical circuit has a source of electrical energy, a path or conductor (copper wire), a control (switch), and some device in which electrical energy is used. Current electricity flows only when the circuit is complete; when the switch is on. The path of electrons from their source through the object they are to operate, and back to their source is called a ***circuit.***

Demonstration

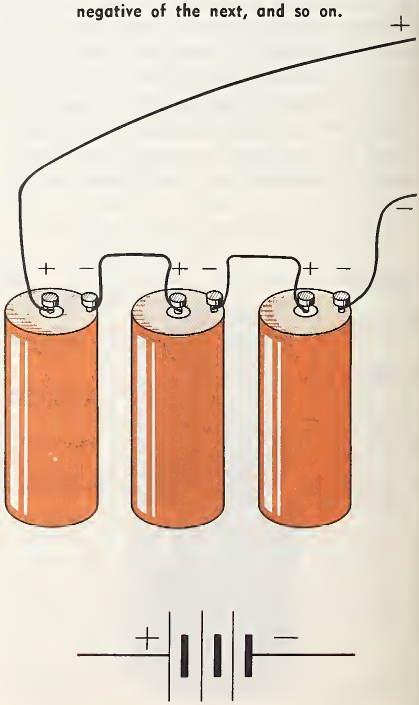
26-2. A simple electric circuit

Set up a simple electric circuit, using a dry cell, some bell wire, a flashlight bulb, a socket, and a knife switch. Close the switch and observe what happens. Can you produce an open or broken circuit in this demonstration? A short circuit? What happens in each case?

A broken wire, an open switch, a burned-out bulb, or a dead cell would result in an open or broken circuit. No current would flow. In the first part of this demonstration, the light went on when the switch was closed. When the circuit was broken, the light went out. A lighting circuit in the home is always a broken circuit until the switch is turned on.

A short circuit will result when the current-carrying wires touch each other. In this case, most of the current does not flow through the complete circuit intended for it, and heat resulting from the short circuit may cause damage. We will see how home circuits are protected from shorts later on.

Fig. 26-4. Cells are connected in series by running a wire from the positive terminal of one cell to the negative of the next, and so on.



Cells in Series. To provide a greater voltage than the 1.5 volts of the single dry cell, we can connect several cells together. This is done by running a wire from the positive terminal of one cell to the negative terminal of the next and so on, as shown in Fig. 26-4. This type of connection is called a series circuit, and the cells are said to be ***in series***. In a series circuit the voltage adds up so that the voltage of the entire series would be the sum of the voltages of the individual cells. What would be the total voltage of two, three, and four dry cells placed in series? Checking with a voltmeter we find that the voltages would be 3, 4½, and 6 volts respectively.

What happens if we add two and three dry cells to the circuit of the flashlight bulb in the previous demonstration? You will note that as cells are added, in series, the bulb burns brighter. The intensity of the light increases as the voltage of the circuit increases. This is because the increased voltage causes more current to flow in the circuit, hence the increase in light intensity. Woodsmen often use flashlights with four, five, and even six dry cells in series to get a strong light beam over a long distance.

Several cells connected in series and enclosed in the same case are called a battery. The battery supplies a higher voltage than a single cell and is portable. Flashlights, as we have said, and portable radios use batteries. In a radio "B" battery, for example, dry cells may be connected in series to provide voltages of 45, 67.5 or 90 volts. Storage cells, in the automobile battery, are connected in series to provide voltages of 6 or 12 volts.

Storage Batteries. Dry cells are cheap and readily portable, but they soon wear out and cannot be recharged. Scientists and engineers finally perfected a longer-lived battery which could be easily recharged and used again. This is called a storage battery. It will last several years with good care. Storage batteries are used in motorboats, automobiles, trains, submarines, airplanes and trucks to supply current for starting, for lights, horn, ignition and many other devices.

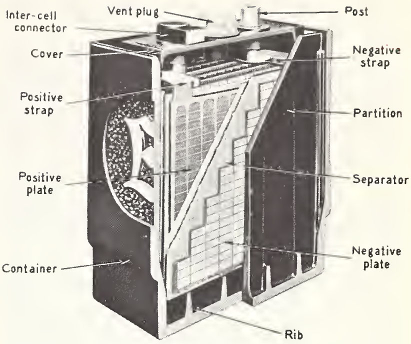


Fig. 26-5. One cell of an automobile battery cut away to show its construction.

What makes up a storage battery? The simple storage battery consists of three storage cells connected in series. Each cell is made up of two types of plates or electrodes, lead (negative) and lead peroxide (positive), immersed in dilute sulfuric acid. These interlocking plates are usually enclosed in a hard rubber case. Wood, plastic, or glass separators insulate the positive from the negative plates. The voltage of a single storage cell is about 2 volts. Modern automobiles use a battery of six cells giving 12 volts.

As the storage battery is used and discharged, chemical action deposits lead sulfate on both plates, reducing the concentration of the sulfuric acid electrolyte. This action weakens the battery, which must be recharged by having direct current passed through it in a direction opposite to that in which the current normally flows. The automobile generator does this when the motor of the car is running. Water lost by evaporation from the cells must be replaced to keep the plates covered. Terminals must be kept clean and the generator kept in good working order.

Electrical Measurement. We have already used two terms of electrical measurement, the volt and the ampere. What is meant by these terms, and what other units of measurement are there?

The ***volt*** is the unit of electromotive force (e.m.f., electrical potential, voltage or pressure) which will cause a current of one ampere to flow through a resistance of one ohm. The ampere is the unit used to measure the rate of flow of electric charge. Current flow involves both quantity and time. The unit of electrical quantity is the ***coulomb***. A coulomb is a quantity of electric charge equal to 6.3 billion times the charge of one electron. The unit of time is usually the second. A rate of flow of charge of one coulomb per second is called a current of one ***ampere***.

The resistance of an electrical conductor to the flow of current is measured in a unit called the ***ohm***. Resistance varies with (1) the kind of material, (2) its cross-sectional area, and (3) the length of the material used as a conductor. Various materials offer different resistances. For instance, silver is a very good conductor with low resistance, while nichrome (a nickel and chrome alloy) is a poor conductor with high resistance. Silver, copper, gold, aluminum, tungsten, nickel, iron and nichrome are listed in order of their increasing resistance to electrical current.

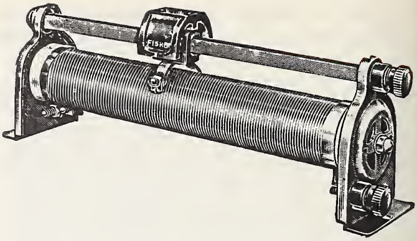
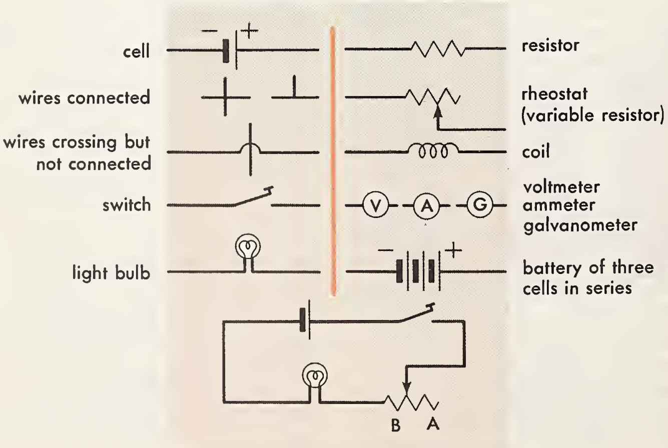


Fig. 26-6. A rheostat is a variable resistor.

The resistance of a wire varies inversely as the cross-sectional area. In other words, the smaller the cross-sectional area, the greater the resistance of the material. For example, one wire having half the cross-sectional area of another wire will have twice the electrical resistance, material and length being the same.

The resistance of a wire varies directly as the length of the wire. A wire 100 feet long will have 10 times the resistance of one of 10 feet of the same material and cross section. Hence, to increase the electrical resistance in a small space, you can use a long, thin wire and wind it into

Fig. 26-7. Notice the standard symbols, some of which are used in a schematic diagram of the circuit discussed in Dem. 26-3.


cell
−
+
wires connected
wires crossing but not connected
switch
light bulb
resistor
rheostat (variable resistor)
coil
voltmeter
ammeter
galvanometer
V
A
G
battery of three cells in series
−
+
B
A

a coil. If we specify the length and cross section of a wire of a given material, we will have a given resistance, no matter what the current flowing through it. Hence an ***ohm*** is best defined as the resistance of a column of mercury 106.3 cm. long, contained in a glass tube of 1 square millimeter cross section.

Resistance offers an opposition to the flow of electric current. Therefore, resistance can be used to cut down the flow of current, or amperes, through an electric circuit. For example, a long, fine nichrome wire may be coiled and a sliding contact may be used on it to control the current flow when placed in series in a circuit. This device is called a ***rheostat.*** (See Fig. 26-6.)

Demonstration

26-3. Varying the current in a circuit

Connect a rheostat in series with the light circuit used in the preceding demonstration, as shown in Fig. 26-7. Close the switch and observe the effect of varying the resistance of the rheostat by moving the sliding contact. Summarize your observations as the contact moves from point *A* to point *B*.

You will note that at point A the current flows through the entire coil of the rheostat, so that the light is dim. The current of the circuit has been decreased. As you move the contact toward point *B*, less of the resistance coil is in the circuit, current flow increases, and the bulb becomes brighter. Recall that increasing the length of a conductor offers increasing resistance and corresponding decreasing flow of current. What happens if the length of the conductor is decreased? There is a definite relationship between current, voltage and resistance. Let us see what this relationship is.

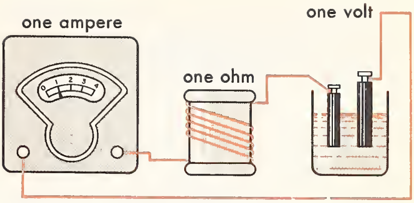


Fig. 26-8. Ohm's law. If a voltage of one volt is applied to a resistance of one ohm, a current of one ampere flows through the circuit.

Ohm's law. We have made some observations in the demonstrations of this chapter; first, that increasing the voltage of an electric circuit increased the current; second, that increasing the resistance, decreased the current. George Simon Ohm, a German scientist, as early as 1826, made the same observations. Ohm discovered the mathematical relationship between voltage, current, and resistance in circuits carrying direct current. We call this relationship ***Ohm's law*** and can state it as follows: ***The current flowing in a conductor is directly proportional to the applied voltage and is inversely proportional to the resistance of the circuit.*** Algebraically written,

$$I = \frac{E}{R}$$

I = current (amperes)

E = applied voltage (volts)

R = total resistance (ohms)

This equation can be more simply written as

$$E = IR$$

In solving a problem using this form, you substitute numerical values for the two known items and perform the arithmetic indicated to find the unknown item. Ohm's law applies to the entire circuit or to any part of it. If you use the formula to determine voltage, current, or resistance of the total circuit, then the values you use must be those of current, voltage or resistance for the entire circuit. When applied to only part of the circuit, the current in that part equals the voltage across the same part divided by the resistance of that part. The units used must be volts, amperes, and ohms.

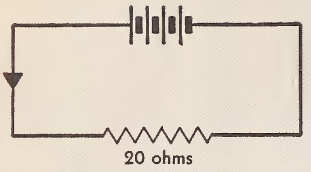


Fig. 26-9. A series circuit.

How would you work a problem using Ohm's law? Let us set up a problem. *Four dry cells are connected in series in a circuit with a resistance of 20 ohms. Find the current flow in the circuit.* (See Fig. 26-9.) You will recall that in a series circuit the voltages add up to give the total voltage. Hence, four dry cells in series, each giving a voltage of 1.5 volts, would have a voltage of 6 volts. Now use Ohm's law:

$$I = \frac{E}{R} = \frac{6 \text{ volts}}{20 \text{ ohms}} = 0.3 \text{ amps}$$

Here is another example. *An electric iron is connected to a 110-volt line. If the iron draws a current of 9 amperes, what is its resistance?* To work this problem we use Ohm's law.

$$E = IR$$

or

$$R = \frac{E}{I}$$

Substitute the values known, E = 110 volts, I = 9 amperes.

$$R = \frac{E}{I} = \frac{110 \text{ volts}}{9 \text{ amps}} = 12.2 \text{ ohms}$$

Measurement of Voltage and Current. How can we easily read the voltage and current of an electrical circuit? There are instruments to measure the electromotive force and the current in an electric circuit. The ***voltmeter*** measures the voltage or the potential difference between any two points of a circuit to which it is connected. The voltmeter is always connected across the circuit or in parallel with the part of the circuit whose potential difference you want to measure, as in Fig. 26-10. There is one precaution that must be observed in connecting meters in a circuit. You must be sure that the positive terminal of the meter is connected to the positive side of the circuit. When these proper connections have been made, the voltage can be read directly from the scale of the meter.

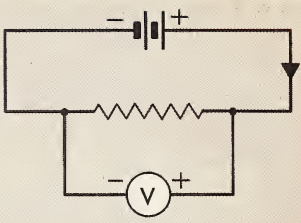
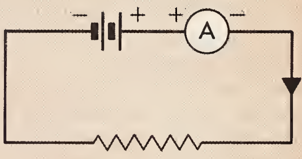


Fig. 26-10. The voltage drop in a resistance is measured by connecting a voltmeter *across* the resistance.

The ***ammeter*** is an instrument used to measure the current flow in an electrical circuit. The current is the same throughout the series circuit, so that in the circuit above, the ammeter can be inserted anywhere in series with the circuit. However, the positive terminal of the ammeter must be connected to the positive side of the circuit. (See Fig. 26-11.)

The next demonstration will show the proper connection of the voltmeter and

Fig. 26-11. Current is measured by placing an ammeter anywhere in series in the circuit.



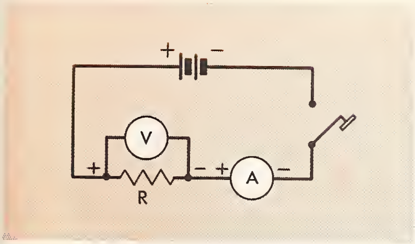


Fig. 26-12. The voltmeter measures the voltage drop across the resistance, and the ammeter measures the current through it.

ammeter and will also give you another opportunity to work a problem using meter readings and Ohm's law.

Demonstration

26-4. A problem in Ohm's law

Fig. 26-12 shows the voltmeter and ammeter properly inserted in a circuit including two dry cells and a flashlight bulb. Take voltage and current readings and find the resistance of the bulb.

V = voltmeter

A = ammeter

R = resistance of bulb

Suppose the meter readings are V = 3 volts and I = 0.3 ampere. Use Ohm's law as follows:

$$E = IR$$

$$R = \frac{E}{I} \quad \frac{\text{reading of voltmeter (volts)}}{\text{reading of ammeter (amperes)}}$$

$$R = \frac{3 \text{ volts}}{0.3 \text{ amperes}}$$

$$R = 10 \text{ ohms}$$

By now you should be proficient in the use of Ohm's law, so let us apply the law to a series circuit of two resistances.

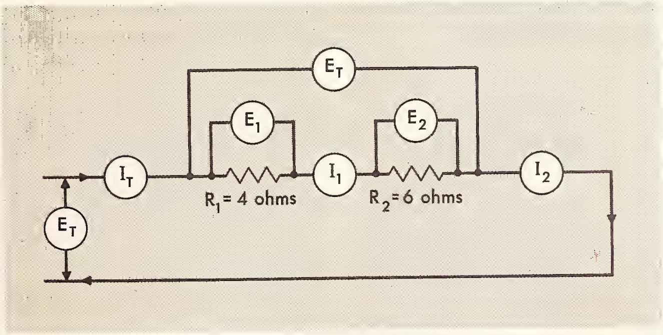
Demonstration

26-5. Voltage, current and resistance in a series circuit

Connect a resistance of 4 ohms in series with one of 6 ohms, using a source of current of four dry cells in series connection. Read the current on either side of the resistances, the voltage across each resistor as well as across both resistors, as in Fig. 26-13. Tabulate your observations. Using Ohm's law, calculate the total resistance of the circuit. Compare the total resistance to that of the two separate resistances. How do the currents compare? What are the voltages across each part of the circuit and what is the total voltage?

To check your work, here are the correct values for the experiment: the current in all parts of the circuit is 0.6 ampere; the voltage across the 4-ohm re-

Fig. 26-13.



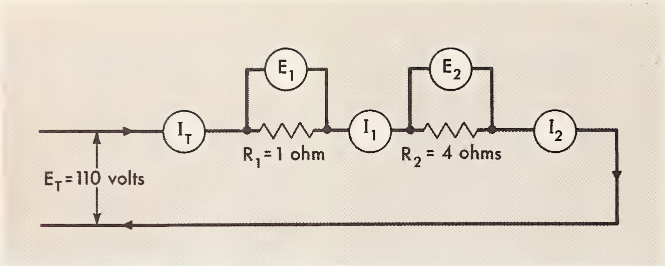


Fig. 26-14.

sistance is 2.4 volts, across the 6-ohm resistance 3.6 volts and the total voltage of the circuit is 6 volts. The total resistance is 10 ohms. The results obtained in this demonstration should show that in the series circuit: (1) the total voltage applied to the circuit is equal to the sum of the voltage drops throughout the circuit; (2) the current stays the same throughout the circuit; and (3) the total resistance of the circuit is equal to the sum of the individual resistances.

Here is another circuit (see Fig. 26-14).

$$E_T = 110 \text{ volts}$$

$$R_1 = 1 \text{ ohm}$$

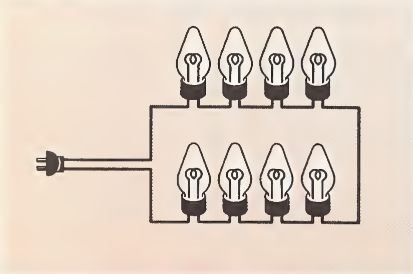
and

$$R_2 = 4 \text{ ohms.}$$

Find R_T, E_1, E_2, I_1, I_2, and I_T.

In your solution use the rule for resistance in a series circuit and find the total resistance. Then using Ohm's law for the entire circuit, calculate the total current. You should know that the current stays the same throughout the circuit (series) so that $I_T = I_1 = I_2$. Using Ohm's law, again, for both parts of the circuit you can calculate E_1 and E_2. Here are the correct answers: $R_T = 5$ ohms, $E_1 = 22$ volts, $E_2 = 88$ volts, $I_T = I_1 = I_2 = 22$ amperes. You can check your results against these. Did you get them all correct?

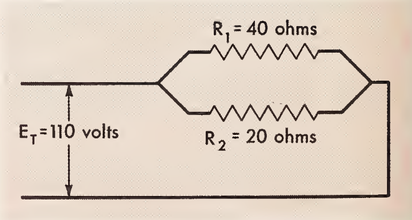
Fig. 26-15. Some Christmas tree circuits, like this one, are connected in series.



A series circuit is not always a good one, as is evident from the old-fashioned Christmas tree string of lights in Fig. 26-15. When one light burns out it breaks the entire circuit and all the other lights go out. This is impractical, and it is also difficult to locate the trouble since each light must be tested separately. Series circuits do, however, have the advantage of increasing resistance to reduce current when this is desirable. Do you remember the rheostat demonstration?

Parallel Circuits. The more up-to-date

Fig. 26-16. The resistances in this circuit are connected in parallel.



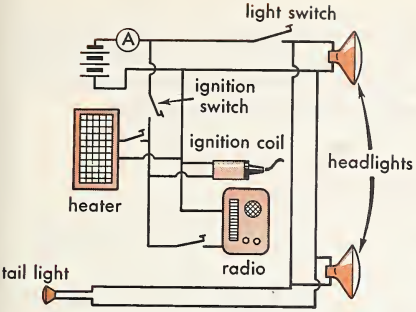


Fig. 26-17. A simplified diagram of an automobile electric system.

Christmas tree light string is an example of another kind of circuit, more widely used, called the *parallel* circuit. In a parallel circuit of dry cells one wire connects up all of the positive terminals and the other all of the negative terminals as in Fig. 26-16 The parallel circuit has the advantage that failure in one part of the circuit will not affect the working of other parts. Home electrical circuits, factory circuits and automobile circuits are good examples of parallel circuits.

Fig. 26-17 shows a common automobile circuit. Lights, radio, heater, and ignition are all connected in parallel with the storage battery of the car. An ammeter is placed between the battery and all other apparatus of the car so that it measures the total current being drawn by them. When the ignition key is turned on and current is drawn by the ignition coil, the ammeter shows a discharge. Likewise, lights, heater and radio cause a discharge. However, when the motor is started and the generator begins to function, the ammeter shows a charge going into the storage battery. The generator, whose power is supplied by the motor, is now supplying all the electric energy needed in addition to recharging the battery.

Parallel circuits obey the following rules: (1) the voltage is the same throughout the circuit; (2) the total current is the sum of the current in each part of the circuit; and (3) the reciprocal of the total resistance is equal to the sum of the reciprocals of the individual resistances that make up the circuit. (The ***reciprocal*** of a resistance is one divided by the resistance.) Or

$$\frac{1}{R_T} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$$

Let us work out an example of a parallel circuit using the rules as stated above and Ohm's law. Fig. 26-16 shows the circuit.

What is the total resistance of a 110-volt circuit consisting of two resistances, one of 40 ohms and the other 20 ohms, placed in parallel? Find the voltage across each resistance and the current flow through each. What is the total current flow?

From the rules for a parallel circuit we find that the voltage is the same, hence the total voltage is the line voltage of 110 volts and the voltage across both resistances is the same, 110 volts. Using Ohm's law for the entire circuit and for branches 1 and 2, we find that $I_1 = 2.75$ amperes, $I_2 = 5.5$ amperes, and $I_T = 8.25$ amperes. The total resistance can be found by using the reciprocal rule of the parallel circuit, whose formula is:

$$\frac{1}{R_T} = \frac{1}{R_1} + \frac{1}{R_2}$$

$$\frac{1}{R_T} = \frac{1}{40} + \frac{1}{20}$$

$$\frac{1}{R_T} = \frac{3}{40}$$

$$R_T = 13.3 \text{ ohms}$$

In the parallel circuit we have provided more paths for the flow of current. For this reason, the total resistance is decreased and more current flows. This fact makes it necessary to protect the parallel circuits used in the home from a current overload when many appliances are used. You will see how this is done in a later chapter.

We have discussed sources of current electricity and the properties of direct current circuits. In later chapters we will discuss the generation of electrical energy at the power plant, its transmission to our homes and factories, and the uses of electricity to produce heat and light.

Things to Remember

There are ***two types of current electricity,*** direct and alternating, abbreviated D. C. and A. C.

In ***direct current,*** the movement of electrons is in one direction, from a negative terminal to a positive one.

In ***alternating current,*** the current flows in alternate directions, rapidly reversing its direction of flow at regular intervals.

In Canada ***60 cycles per second*** is the most common alternating current.

An ***electric circuit*** is a complete path for electricity.

The strength of an electric current is measured in ***amperes.***

Electromotive force or voltage is measured in ***volts.***

Voltage may be increased by connecting several cells in series to make a battery.

Electrical resistance is measured in ***ohms.***

The ***resistance*** of a wire increases as its length increases.

The resistance of a wire decreases as the area of its cross section increases.

Ohm's law states:

$$\text{Current (amperes)} = \frac{\text{Voltage (volts)}}{\text{Resistance (ohms)}} \quad \text{or} \quad I = \frac{E}{R}$$

Questions

GROUP A

1. What is the symbol for the following: (*a*) cell, (*b*) resistor, (*c*) voltmeter, (*d*) ammeter, and (*e*) cells in series.
2. Show how to connect an ammeter and a voltmeter in an electric circuit.
3. What is meant by the terms (*a*) ampere, (*b*) volt, and (*c*) ohm?
4. Make a drawing of a dry cell and label its parts.
5. List the rules for (*a*) a series circuit (*b*) a parallel circuit.
6. Set up a series circuit and tell how to calculate all resistances, voltages and currents, using Ohm's law.
7. What current flows through a 12-ohm flashlight bulb when used in a two-cell flashlight?
8. What is the resistance of a 110-volt light bulb that draws 0.55 amperes?
9. How many volts must be applied to a 300-ohm resistor to cause a current of 0.2 amperes to flow through it?
10. What is the total resistance of a 4-ohm resistance and a 6-ohm resistance connected in parallel?

GROUP B

11. Calculate the total resistance of three resistors of 5, 10, and 15 ohms connected in parallel to a 110-volt line. What is the voltage across each, and the current flow in each? What is the total current?

12. Wire up three electric lights in parallel and measure the voltage and current, then calculate the resistances and the total resistance.

13. List the advantages of series and parallel circuits.

14. Five dry cells in series are in a flashlight. The current flow is 0.25 ampere, what is the resistance of the light bulb?

15. What is the total resistance of five 10-ohm resistors connected in parallel?

16. A 3-ohm and an 8-ohm resistor are connected in series to a 5.5 volt battery. (*a*) What current flows through each resistor? (*b*) What is the voltage drop across each resistor?

17. The resistors in problem 16 are connected in parallel to a 1.5 volt cell. What is the current through each?

Things to Do

Switch circuits • Connect two single-pole, double-throw switches with a battery and light bulb as in Fig. 26-18. This circuit is the kind that is used to control a stairway light so that the light may be turned on and off from either the bottom or the top of the stairs. Each switch controls the light independently of the position of the other. Demonstrate the project to your classmates. If possible, obtain a suitable switch, the kind used in actual practice in your home. Study the construction of this switch. It works, in principle, the same way as the single-pole, double-throw switch you used. (These switches are always used in pairs.)

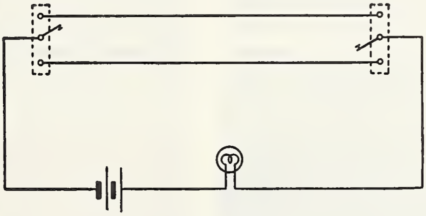


Fig. 26-18. Double-throw switches.

27

HEAT AND LIGHT FROM ELECTRICITY

Just imagine a home without an electric light, an electric iron or a toaster. Would you want to go back to trimming lamp wicks and cleaning lamp chimneys? Perhaps you would enjoy toasting bread in the oven of the gas stove or heating an old-fashioned iron on the stove to press your clothes. Electric lights, irons, percolators, stoves, toasters, heaters, and the electric waffle iron are taken for granted. We expect to have them ready to work and at our finger tips to use.

We watch football and baseball games under strong batteries of lights, something unheard of in the last generation. Then we come home, settle down in our favorite chair, turn on the lamp and comfortably enjoy the newspaper or the latest magazine, listen to the radio, watch TV, or start to study. Every modern home uses electrical energy.

How is it possible to get light and heat from an electric current? We have learned, previously, that resistance is the opposition to the flow of electrons through a conductor and that electrical resistance depends upon (1) the length of the conductor, (2) the cross-sectional area of the conductor, and (3) the kind of material of the conductor. Let us examine the nichrome heating element of an electric heater. Here we find a thin wire of a nickel-chromium alloy coiled up so as to gain length. Nichrome is a material of high electrical resistance. So we have all three of the resistance characteristics satisfied. Let us place the heating element in a socket and connect it to a 110-volt line. We notice that the wire gets quite hot and gives off a lot of heat, and a red glow of light at the same time. Take a 150-watt electric bulb and place it in a socket similar to the one used previously. Connect to the 110-volt line and observe. This time the tungsten filament of the bulb gets white hot, or ***incandescent.*** Heat and light are both observed. Incandescence is always accompanied by intense heat and light.

Why does the filament get hot? The copper wire which brings the electric current to the bulb is only slightly warm. The filament, however, is different from the copper wire in two ways—in size, and in the material of which it is made. These factors create a high resistance to the flow of electric current. The greater the resistance, the more difficult it is for the electricity to pass through the tungsten filament. What effect does this have on the heat produced? To answer this question we can perform an experiment.

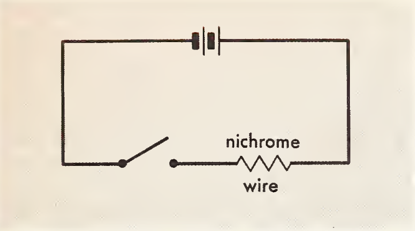


Fig. 27-1. A short piece of nichrome wire becomes incandescent when connected to two dry cells.

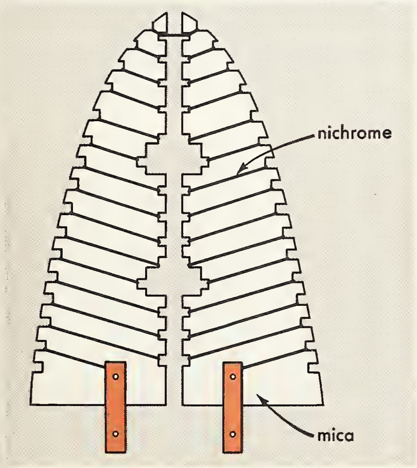
Demonstration

27-1. The heating effect of an electric current

Connect a piece of fine nichrome wire about four inches long to two dry cells and a switch as shown in the diagram. Use copper wire of the same cross section to complete the circuit, and then close the switch. The nichrome wire becomes so hot that it glows, the copper does not. Vary the length of the nichrome wire. What do you observe? If you have some iron wire of the same cross section, try an equal length of it. Observe what happens. Can you summarize your results?

You have seen nichrome wire has a high resistance, and that the heating effect depends on the substance as well as on the size and length of the resistor. Thus nichrome is a good wire to use in electric heaters, toasters, and irons. In the electric iron, nichrome wire is wound around a mica core; in the electric heater around a porcelain core. The heater may have a fan behind the heating element, so that the air is blown past the heating coil and strong convection currents are set up in the room.

Fig. 27-2. The heating element of an electric iron.



Electrical appliances are easy to use, they are clean, and do not produce objectionable gases. A flick of the switch sets them into operation. Electric heaters for room heating are usually used for auxiliary or emergency purposes. However, in some localities, hot water is heated by means of electricity. Here the heating takes place primarily during the hours when peak loads of electrical energy are not required, and consequently the cost of the electrical energy used is at a lower rate.

Electric stoves, hot plates, percolators, clothes dryers and soldering irons are

Fig. 27-3. The familiar electric stove. This stove has four heating units.

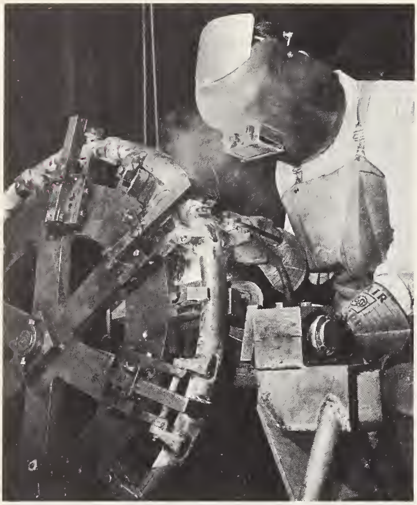


Fig. 27-4. Electric welding. A circuit is completed by making contact between two metal rods. The high resistance at the point of contact causes the tips of the rods to melt and fuse together.

among other heat-producing appliances, convenient to operate from electricity. The electric blanket is also becoming very popular since it is warm and has little weight. Electric resistance methods are used to heat the flying suits of aviators and to de-ice the wings of planes. Industry uses electrically produced heat in drying ovens, electric welding and in electric furnaces.

In electric welding, the electric current flows through the two metal parts to be welded together. The point of contact is the place of highest resistance, hence the metals actually fuse here, forming a strong bond. This welding process is used as a substitute for the former method of joining by rivets, in joining small parts or in constructing bridges, ships and buildings.

In the electric furnace, the material is melted either by the temperature of an electric arc between two electrodes or by passing the current directly through the material itself. Here the resistance of the contents is sufficient to produce the high temperature needed. The electric furnace is used to manufacture graphite from carbon (coke), to refine aluminum and produce high-grade steel.

Light from Electricity. Thomas Alva Edison developed the first practical electric lamp, using a carbonized cotton thread in an evacuated glass bulb. Tungsten wire is used today, as we have already mentioned. Nichrome wire cannot be used, since it melts at a temperature below the white-hot level necessary for the most efficient production of light. Tungsten's melting point is above 3000° C, so that it can easily be heated to incandescence without melting by the passage of an electric current. However, the filament must be long and thin. Wire leads carry electricity to the tungsten filament. These leads are sealed into the glass bottom of the bulb and in turn the entire glass bulb is cemented into a screw plug.

Multiple filaments are used in some cases so that several different amounts of light may be had from the same light bulb, as in the so-called three-way bulb. For example, there is the 50-100-150 watt bulb. The first time you turn the switch one filament glows, giving the light of the 50 watts; a turn of the switch turns off this filament and sends current through another filament to give the light of the 100 watts; the last turn of the switch throws both filaments in parallel so that the light gets brighter, operating on the full 150 watts. Similar bulbs come in larger sizes as in the 100-200-300 watt size.

Fluorescent tube lighting is used in many industries, schools and offices. This lighting is popular in some parts of the modern home—the kitchen, the bathroom and in lighting in dining and living areas. Fluorescent tubes are very efficient, use little power, are fairly cool, and their cost of operation is small, although their initial cost is high.

What is a fluorescent tube and how does it give light? A glass tube is coated on the inside with a powder which glows under ultraviolet radiation. The ultra-

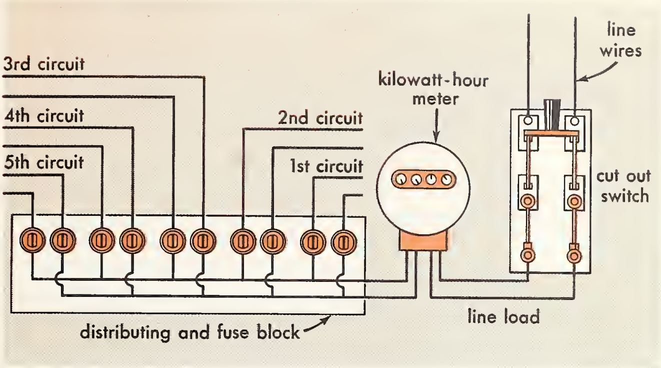


Fig. 27-5. A house circuit. Notice that all the current entering the house may be cut off by a main switch. Notice also that the current is first passed through a meter before it is distributed along different circuits.

violet rays come from mercury vapor resulting from the evaporation of a small amount of mercury that is in the tube along with some argon gas. A special starter switch along with a ballast or choke coil causes a mercury vapor discharge between the electrodes in the tube. This discharge is rich in ultraviolet rays that impinge upon the fluorescent coating, causing it to give off an intense glareless light restful to the eyes. The color of the light can be changed by coating the tube with various chemical salts. A fluorescent light is cheaper to operate than an incandescent tungsten light; for example, a 40-watt fluorescent tube will produce about the same amount of light as a standard 150-watt incandescent bulb.

The ***neon lamp,*** which we have all seen, is another tube using electricity to produce light. Neon, one of the rare gases in the air, glows with a very noticeable luminescence. Neon tubes are made in many shapes and designs; different colors can be produced by using different gases. Neon alone produces the red color you have seen. Mercury will produce blue-white. Neon tubes are especially useful in making signs and decorations. Transformers are required to boost the voltage from the usual 110 volts to 10,000-12,000 volts to operate a neon tube. The cost of operation is quite small, as you might expect from seeing many store neon signs in operation both day and night.

Sodium vapor lamps are very useful for highway lighting and for bridges. They consist of glass tubes with tungsten filaments, a small piece of pure sodium, and some neon gas. Electricity heats the filaments which in turn vaporizes the sodium. This vapor conducts the electric current, giving off a soft, yellow light. Visibility is better under a yellow light, so sodium vapor lamps are used at dangerous intersections, curves and on bridges.

The House Circuit. How does electricity get to the appliances it is to operate? Let us look at a house elecrical circuit diagram, Fig. 27-5. Electricity comes into our home by means of lead-in wires. Usually three wires enter the house. The voltage across either of the outside wires and the center wire will be about 110 volts. The voltage across the two outside wires will be about 220 volts, which is needed to operate some pieces of electrical equipment such as an electric stove. The center wire of the three is grounded. Fig. 27-6 shows a two-wire circuit which is a simpler circuit sometimes used. The voltage of this line will be 110 volts. It

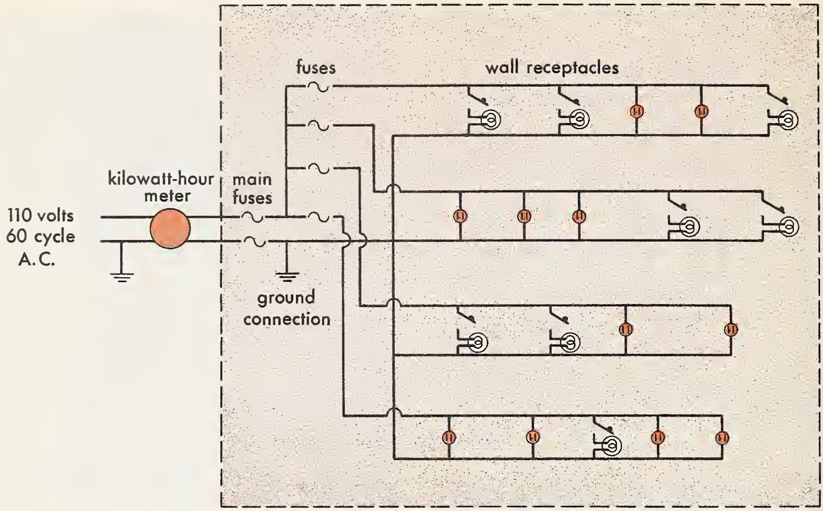


Fig. 27-6. This simple house-wiring system shows four different circuits.

is 60-cycle A. C. electricity. The wires lead in through a kilowatt-hour meter which measures the amount of electrical energy we use. From here a wire leads through the main fuse, then goes through the individual circuit fuses and then to the individual circuits of the house; the other as a common wire leads to the other side of each electrical outlet. The diagram here shows four separate circuits, three having five outlets, and the other four. Note that all are connected in parallel so that each part of the circuit and each circuit is independent of the others. Switches are shown for lamps or ceiling lights.

Fuses. The fuses in the circuits act as safety devices to prevent overloading of the circuit. Separate fused circuits insure that if one fuse goes out, electrical appliances on other circuits are not disturbed. Usually the main fuses are 50-ampere fuses and the branch fuses are 15-ampere. These ratings mean that the fuse will blow if more current than the ampere rating is drawn.

What is a fuse? We have said that a fuse is a safety device to prevent the overloading of an electric circuit. We have seen that current causes heat in a conductor and that wires can get very hot, especially if the current is large. A hot

Fig. 27-7. Two kinds of fuses.

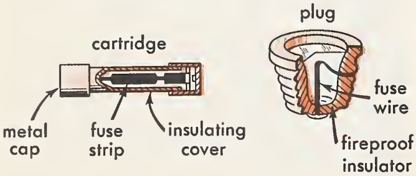
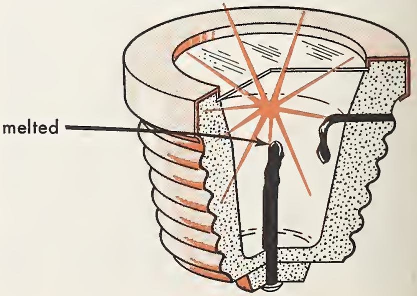


Fig. 27-8. If the current is too large, the fuse melts and the circuit is broken.



wire inside a wall or near any combustible material could very easily start a fire. Copper is an excellent conductor and when connected in a working circuit with some high-resistance appliance, functions well. However, if a short circuit results, so that only the copper wire of low resistance is involved, then a very high current flows, causing tremendous heat. A fuse is a short piece of lead metal alloy with a low melting point. This strip is mounted in a screw-type holder, so that it screws into an electrical socket and can be placed in series with the circuit. The holder may have a mica window at the top so that the fuse wire (a flat strip) is visible. If a current larger than the current rating of the fuse passes through it, the fuse metal alloy is melted, breaks the circuit and turns off the electricity. Be sure to check all parts of the circuit in question before replacing a burned-out fuse. Fuses should never be replaced with fuses of higher rating.

Some fuses are of a cartridge type where the fuse metal is placed inside a tube with metal ends which fit into prong holders in the fuse box. The cartridge fuse is usually used in circuits carrying large currents. Modern homes may have a circuit breaker. This device, an automatic switch operated by electromagnets, automatically opens a switch in the circuit when the wires become overloaded.

Demonstration

27-2. Blowing a fuse

Take a piece of metal foil from a candy-bar wrapper and cut a strip about 2 inches long and ¼ inch wide. At the center cut a V-shaped notch from each edge so that there is a very thin spot at this position of the foil. Secure the strip with two thumb tacks to a small wood block as shown in Fig. 27-9. Using a circuit with a dry cell, a knife switch and a flashlight bulb, connect your improvised fuse in series with the bulb. Throw the knife switch and observe what happens. Hold a short piece of insulated wire across the two terminals of the lamp socket and observe the results. What have you done?

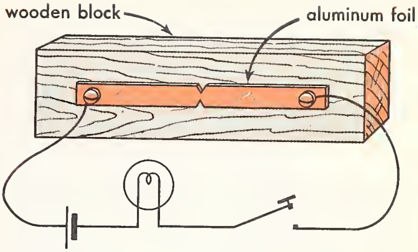


Fig. 27-9. Aluminum foil can be used as a fuse wire.

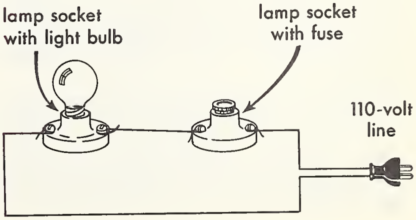
When the knife switch was first thrown the bulb lighted since current from the dry cell was passing through the metal foil and to the bulb, whose resistance was such that the current to light the bulb was not enough to blow the fuse. However, when you touched the metal wire across the terminals of the bulb you produced a short circuit (a low-resistance path) for the current. The heavy current flowing in the circuit passed through the fuse, melting the metal foil. This shut off the flow of electrons (current), and the circuit was broken. This demonstration indicates the safety value of a fuse. An alternate demonstration is as follows:

Demonstration

27-3. Blowing a fuse

Use two standard light sockets and wire them in series as shown in Fig. 27-10. Place

Fig. 27-10. The fuse wire melts if the lamp is short circuited. Why?



a 10-ampere fuse in the first socket and a 100-watt bulb in the second. Attach the circuit to a 110-volt line. Observe what happens. Now touch a heavy, *insulated* wire across the terminals of the lamp socket. What happens?

When the circuit is attached to the 110-volt line the bulb lights. The resistance of the bulb is such that less than one ampere of current flows through the circuit and of course through the fuse. Remember that this fuse is rated at 10 amperes. However, when we short circuit the lamp with the heavy, low-resistance wire across its terminals, a large current flows through the fuse, the fuse wire melts and the circuit is broken.

Safety Measures. There are a number of safety measures to consider in the home. Be sure to correct the cause of the blown fuse before replacing it. Always use standard fuses of proper rating. If an electric circuit in your home needs repairing, always employ a competent electrician, and do not try to do the work yourself. Never handle electrical appliances when your hands are wet or your feet are in contact with a wet, grounded surface such as a bathtub, or the basement floor. Electric fixtures in the bathroom should not be placed within reach of the bathtub. Keep all appliances in good working order.

Voltage and Power Ratings. Electric appliances are usually rated according to the line voltage needed and power consumed in watts. These figures are found on the plate attached to the appliance or on the top of a light bulb. For example, a certain electric iron is marked 120 volts, 1000-watts; an electric light is marked 120 volts and 100 watts; a toaster is marked 110-120 volts and 1100 watts. Sometimes the current rating is given in amperes.

What do these ratings mean? Most home voltages are 110 to 120 volts. We usually use the 110-volt figure in electrical problems, unless we actually measure the voltage of the circuit involved. The watt is the unit of electric power. ***A watt is the power required to keep a current of one ampere flowing in a circuit under a pressure of one volt.*** This may be expressed in a relation between watts, amperes and volts, as follows:

$$\text{Electric power} = \text{voltage} \times \text{current}$$

$$P = EI$$

We can use this equation to find out how much power is consumed by an electric appliance if we know how much current it draws and the voltage it operates on. For example: *How much power is used by an electric toaster that draws 10.5 amperes from a 120-volt line?* Since we know the voltage and current, we can substitute in the above equation for electric power:

$$P = EI$$

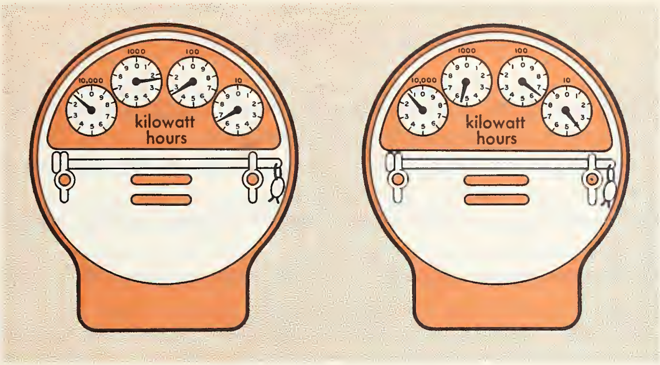
$$P = 120 \times 10.5$$

$$P = 1260 \text{ watts}$$

The watt is a rather small unit of power, so we often use the ***kilowatt,*** which is 1000 watts. The above toaster, for example, consumes 1260 watts or

$$\frac{1260}{1000} = 1.26 \text{ kw.}$$

Electric Energy. In studying simple machines such as pulleys and inclined planes, you learn that power is the rate at which work is done. Hence, if we know the power of a machine and the length of time it does work, the total work done will be the power times the time it operates. Electric power and energy are fundamentally the same as mechanical power and energy and obey the same laws. It therefore follows that the electric energy consumed by an appliance is equal to the power it consumes times the time it operates. In a circuit whose voltage is one volt, a current flow of one ampere for a time of one hour uses one watt-hour of electric energy or work. Again, since the watt-hour is a small unit, electric work or energy is usu-



meter reading at start of month

meter reading at end of month

Fig. 27-11. A kilowatt-hour meter. At the beginning of the month the meter reads 1236; at the end it reads 1564. Hence 328 kilowatt-hours of electrical energy were used during the month.

ally measured in ***kilowatt-hours;*** one kilowatt-hour being equal to 1000 watt-hours. In equation form:

Electric energy = electric power × time

Energy (watt-hours)
 = E (volts) × I (amperes) × T (hours)

or

$$\text{Energy} = \frac{EIT}{1000} \text{ Kilowatt-hours}$$

A kilowatt-hour meter is installed in your home to measure the amount of electric energy that you use in your home circuits. This meter is read once a month by the meter reader from your power and light company. The difference between two monthly readings indicates the number of kilowatt-hours used and when multiplied by the cost per k.w.h. indicates the amount you pay for this electric energy. Fig. 27-11 shows the kilowatt-hour meter. Notice the four hands and that some rotate clockwise and others counterclockwise. This is due to the gear train of the meter, whereby the mechanism of the meter uses gear wheels that are meshed with one another. As one turns clockwise, the ones in mesh with it obviously turn counterclockwise. You can read the meter by starting with the dial on the left and reading from left to right, always reading the number that the hand has just gone by.

Here is a problem. Can you do it? *The readings of an electric meter taken a month apart are 3302 k.w.h. and 3217 k.w.h. If electricity sells at five cents a kilowatt-hour, what would you pay for the electrical energy you used?* Did you get the answer? It is $4.25, obtained by finding the difference between the two readings, 85 k.w.h., and multiplying this by the $0.05 per k.w.h. cost.

Let us do another problem involving cost. *Your mother is ironing. She uses a 110-volt iron whose current rating is 10 amperes, for one half-hour. What was the cost of operation at rate of $0.05 per k.w.h.?* The answer is $0.0275. Using the equation:

$$\text{Energy} = \frac{EIT}{1000} \text{ k.w.h.}$$

$$\text{Energy} = \frac{110 \times 10 \times 0.5}{1000} = 0.55 \text{ k.w.h.}$$

The cost is 0.55 k.w.h. × $0.05 per k.w.h.
 = $0.0275

Things to Remember

Heat is produced when an electric current flows through a resistance.

Nichrome wire is used in ***electric heating elements.***

Electric heat is cleaner, safer, and more convenient to control than heat produced directly by combustion.

An incandescent bulb contains a fine ***tungsten filament*** because tungsten has a very high melting point.

Fluorescent tubes operate at lower cost and with less heat than incandescent bulbs.

House lights and appliances are ***wired in parallel,*** and each circuit is protected by a fuse.

An ***electric fuse*** is a lead alloy with a low melting point that melts when an excessive current flows through it, and thus breaks the electric circuit.

The ***watt*** is the unit of electrical power.

Electrical energy is purchased by the ***kilowatt-hour.***

Questions

GROUP A

1. What kind of wire is used in electric heaters? Why is this kind of wire used?
2. Give three advantages of electric heating over fuel heating.
3. What is an electric fuse, and what does it do in an electric circuit?
4. Describe the operation of a sodium-vapor lamp.
5. List four safety rules for handling electricity.
6. List the power rating, the voltage rating, and the current rating of three home electrical appliances. Calculate any of these not directly given.
7. How is a fuse connected in an electrical circuit? Why?
8. What is a short circuit?
9. Why is a fuse not connected in parallel with a circuit?
10. Why is heat produced from electricity more expensive than heat produced directly from burning coal?
11. Explain the operation of a fluorescent tube.

GROUP B

12. How much power is used by a 110-volt light bulb that draws a current of 0.5 amperes?
13. What is the resistance of the light bulb in the preceding problem?
14. What current flows in a one kilowatt, 120-volt electric toaster?
15. How much electric energy is used by the toaster in the preceding problem in making a piece of toast, if it takes two minutes to make the toast?
16. What is the cost of making toast with the toaster in problem fourteen if electric energy costs 5 cents per kilowatt-hour?

Things to Do

To make an arc light · Rig up a stand to hold two carbon rods mounted vertically and not touching. (Carbon sticks from flashlight cells will do.) Wire a nichrome heating coil in series with one of the carbons. Connect your lead-in wires to a plug to connect to the 110-volt line. (One lead-in wire goes to the heating coil, mounted in a socket, and from the other socket terminal to the lower carbon rod. The other lead-in wire goes from the plug to the top carbon.) To operate bring one of the carbons in contact with the other and after they are heated, move slightly apart so as to cause an arc. A bright light will result. ***Caution:*** Do not look directly at the light, it may give you a headache; do not touch both carbons at the same time when adjusting them, or you will probably get a shock.

28

POWER FROM ELECTRICITY

WE HAVE SEEN that dry and wet cells are used as portable power plants to provide direct current. You will remember that the chemicals of a dry cell are eventually used up and that wet cells, especially the storage battery, have to be recharged regularly. Are there other ways of producing electric power?

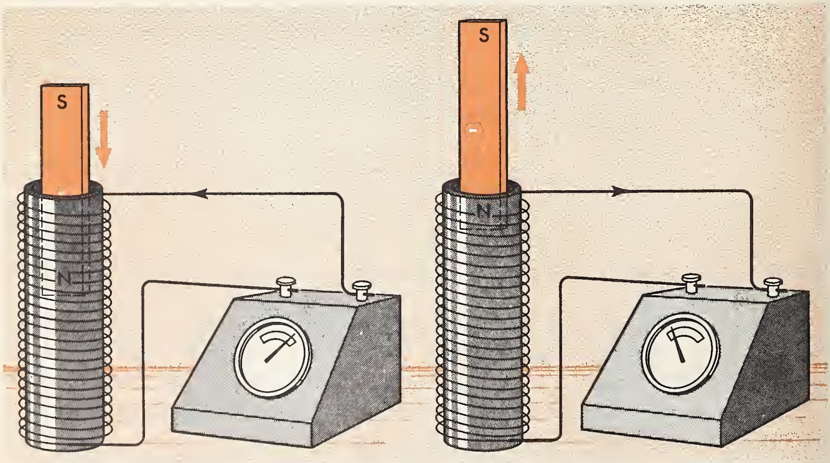
Oersted, the Danish physicist, in 1819 showed that a wire carrying an electric current was surrounded by a magnetic field. Faraday, an English scientist, wondered about the reverse effect. Could magnetism be used to induce an electric current in a circuit? He used a simple apparatus, as in the following demonstration, to test his theory.

Demonstration

28-1. Generating an electric current in a coil by means of a moving magnet

Take a cardboard mailing tube about six

Fig. 28-1. A current is generated by thrusting a bar magnet into and out of the coil.



inches long and wind several hundred turns of fine insulated magnet wire around it. You can keep the wire in position with tape or by shellacking it. Connect the ends of the coil thus formed to the terminals of a galvanometer. Quickly thrust one end of a bar magnet in and out of the hollow coil. (See Fig. 28-1.) What happens? Hold the magnet stationary inside the coil. Observe the galvanometer when the magnet is (1) moving in the coil; (2) moving out; and (3) when the magnet is stationary within the coil. When is current generated? Does it make any difference if the coil moves and the magnet remains in a fixed position? Try it.

You will notice that when the magnet is moving into the coil the galvanometer needle is deflected, indicating that an electric current is being produced. When the magnet is moving out of the coil the galvanometer is again deflected, but in the opposite direction. However, when the magnet is stationary, inside the coil, no current is produced and the galvanometer needle remains at zero position.

To summarize the results of this demonstration we may say that a current of electricity is induced in a coil when the magnetic field in the coil is changing, and that the direction of the current depends upon the motion of the magnet or coil. The intensity of the current depends upon the rapidity of the movement.

Three conditions are necessary to generate an electric current: (1) a ***magnetic field,*** here supplied by the bar magnet; (2) a ***conductor,*** here supplied by the coil; and (3) ***motion,*** here supplied by your hand. These three things are all that are necessary in any generator or dynamo.

Generators. What is an electric generator? ***An electric generator is a machine that converts mechanical energy into electrical energy.*** It has two main parts, an armature and a magnetic field. The armature consists of many coils of insulated wire supported on a shaft that revolves in a magnetic field.

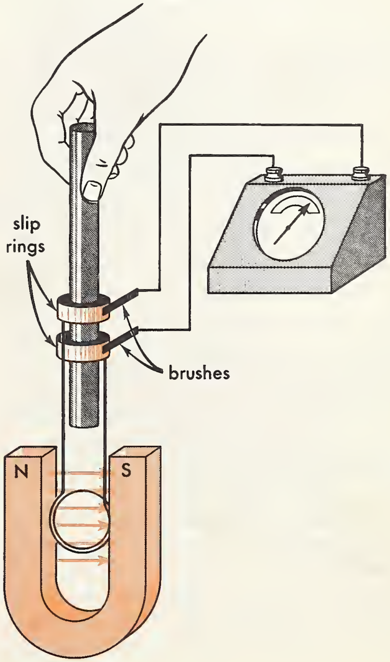
Demonstration

28-2. A simple A. C. generator

Mount a strong horseshoe magnet vertically on a block of wood. Mount two wooden washers about one inch in diameter and one inch apart on a ¼ inch wooden dowel rod. Use metal foil and make rims for the washers. Connect several coils of wire as shown in the figure so that one end connects to slip ring #1 and the other to slip ring #2. Use your ingenuity and mount the rod so that the coils can turn across the lines of force of the magnetic field of the horseshoe magnet. Connect brushes, of tin or thin copper, to the slip rings and wire them to a galvanometer. (See Fig. 28-2) Turn the shaft slowly and then more rapidly. What happens? Explain.

In this demonstration you have provided the necessary requirements for a

Fig. 28-2. A simple A.C. generator. If a coil is rotated between the poles of the magnet, an alternating current is generated.



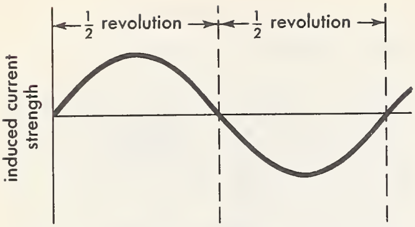


Fig. 28-3. For each revolution of the coil the current increases from zero to a maximum, then back to zero; then it reverses. This is an alternating current.

generator. Note that as you turn the shaft, the coils of the armature cut across the lines of force of the magnetic field of the magnet, thus inducing a flow of current in the armature coils. The galvanometer needle will move in one direction, then back to zero, and in the other direction and back to zero. This is repeated, indicating the presence of an alternating current of electricity. (See Fig. 28-3.)

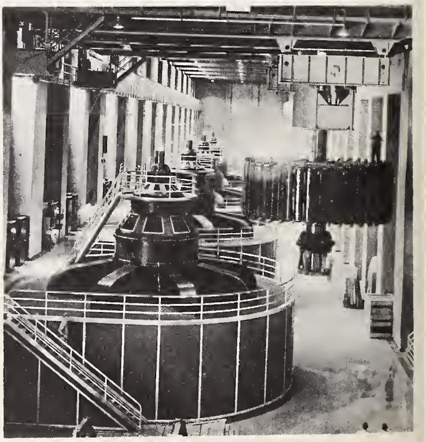
Commercial Generators. In large generators the current is usually induced in fixed coils on the frame, called the ***stator,*** while the rotating field, whose lines of force cut these coils, is produced by sending a small current through the brushes to electromagnets mounted on the rotating shaft, called the ***rotor.*** This arrangement is followed because the field current in the rotor need only be very small, causing little sparking and heating at the brushes. A small generator called an ***exciter*** furnishes current for the rotor. The current induced in the fixed coils, on the other hand, represents the total current output of the generator. In a large generator this current is so great that if it were induced in the rotor it would produce excessive sparking and heat at the brushes, which would be rapidly worn away.

Powering the Generator. How are commercial generators powered? Commercial generators may be powered by steam or by water. Steam is generated from the energy of burning coal. Coal is usually finely powdered and blown with the proper amount of air into the burner where it is ignited. The coal burns almost completely with little loss in ash. The heat of the burning coal converts water in the boiler to steam which, in turn, drives a turbine. The rotating shaft of the steam-driven turbine is connected directly to the rotating shaft of the electric generator. An electric current is induced in the commercial generator, just as it was in the demonstrations described in this chapter. Spent steam from the turbine is used to heat the power house and is then condensed into water which is used again in the boilers.

Where water is plentiful and a good head can be supplied, generators are turned by the power of falling water channeled against the turbine blades. Such places as Hoover Dam, Niagara Falls and Norris Dam are examples of hydroelectric power.

Alternating Current. The current generated by the type of generator shown in Fig. 28-2 is an ***alternating current;*** that

Fig. 28-4. Four 16,500-volt generators at Hoover Dam. Each produces about 80,000 kilowatts of electric power. The shafts of these generators are mounted vertically and are connected to huge water turbines beneath the floor. Note the large revolving field from one of these generators suspended from a crane.



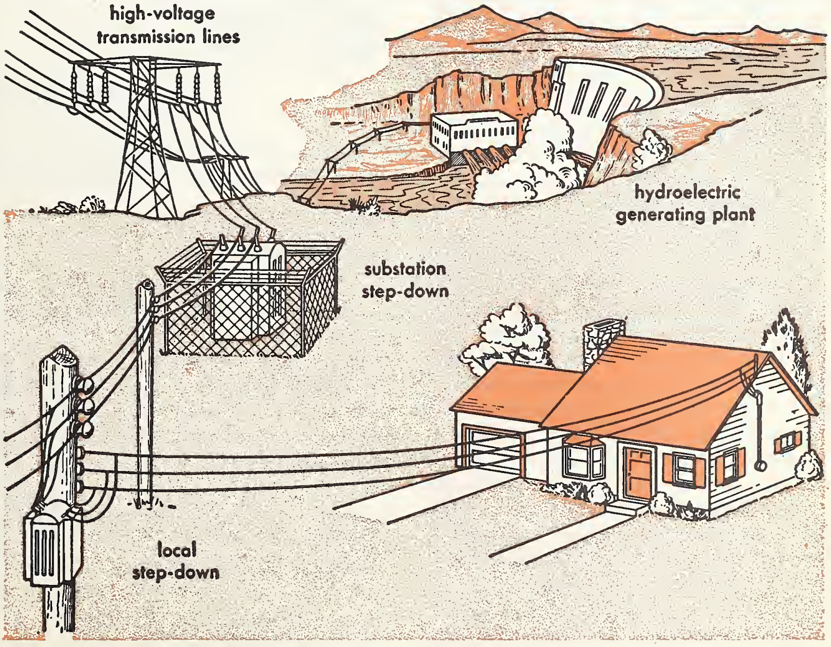
is, a current which flows first in one direction, and then reverses and flows in the opposite direction, as shown in Fig. 28-3. In the large alternating-current generators, called ***alternators,*** this process is repeated 120 times a second. To see why such generators produce an alternating current, we will recall that in Demonstration 28-2 the direction of the induced current depended on the direction of the motion which caused the current. The revolving rotor of a generator has the same effect; that is, the direction of motion of the wires in a single loop of the armature that cuts the field produced by the stator, changes every half-revolution as shown in Fig. 28-3.

Sixty-cycle alternating current is used in home circuits. Why do we use alternating current? The electrical energy we use, as we have already seen, is measured in kilowatt-hours. The units ***watts*** or ***kilowatts*** involve the two factors, voltage and current. Current produces heat in an electrical conductor; the greater the current the greater the heat. If there are many miles of wire between the power plant and our homes a great deal of electrical energy is lost in the form of heat, particularly if the current is large.

We can prevent this great heat loss if we step up the voltage (electrical pressure), thus cutting down the current value over the long transmission. If we can reverse this procedure near the consumer so as to provide a high current at a lower voltage, this would be an ideal situation.

The transmission of electrical energy from the power plant might be like this: generator (13,200 volts) → transformer bank (132,000 volts) → substation trans-

Fig. 28-5. Transmission of electric power from a generating station to the consumer. Transformers step up the voltage for line transmission and step it down before electricity can be safely used in houses.


high-voltage transmission lines
hydroelectric generating plant
substation step-down
local step-down

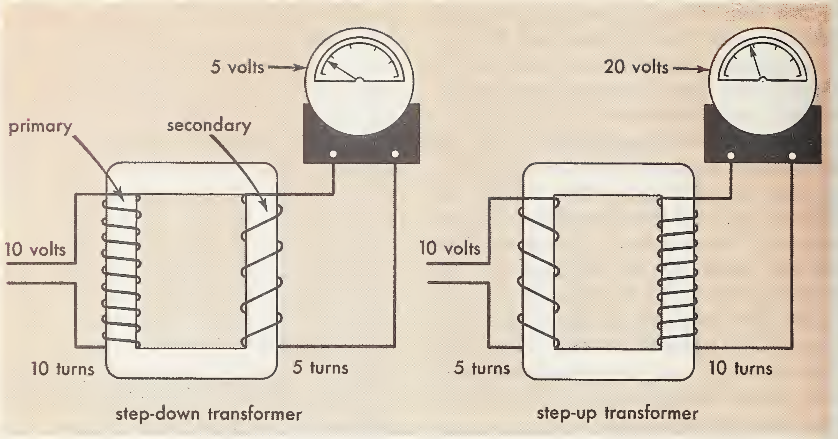


Fig. 28-6. A transformer is a voltage changer.

former (66,000 volts) → feeder substation transformer (13,200 volts) → feeder transformer (4400 volts) → pole transformer (220) → 3-wire system to home (220 volts and 110 volts). As the voltage goes up the current goes down and vice versa. For longer transmission distances the voltage may be increased to 100,000 or even to 300,000 volts.

Alternating-current electricity is used for most circuits in industry and in our homes because its voltage can be boosted up and cut down easily with low energy loss between producer and consumer.

Transformers. How is the stepping-up and stepping-down of voltage accomplished? This is done by means of a ***transformer,*** which consists of an iron core on which two coils are wound. One coil is called the primary, the other the secondary. The coils are insulated from each other. A varying magnetic field in the iron core produced by an alternating current in the primary coil induces a voltage in the secondary coil.

If the number of turns of wire in the secondary coil is greater than the number in the primary, the voltage is increased. This is a ***step-up*** transformer. If the number of turns on the secondary coil is fewer than those of the primary, the voltage is decreased. This is a ***step-down*** transformer. Fig. 28-6 shows the construction of a transformer.

A mathematical equation which expresses the relationship between the turns of the primary and secondary coils and the voltages in each is

$$\frac{\text{Voltage in primary}}{\text{Voltage in secondary}} = \frac{\text{Turns on primary coil}}{\text{Turns on secondary coil}}$$

In the mathematical equation between current and turns in both coils we have

$$\frac{\text{Current in secondary}}{\text{Current in primary}} = \frac{\text{Turns on primary coil}}{\text{Turns on secondary coil}}$$

Let us take an example: ***A transformer has 15,000 turns on the primary and 500 turns on the secondary. If 3300 volts are applied to the primary, what is the voltage of the secondary?***

$$\frac{3300}{\text{Volts on secondary (or E)}} = \frac{15{,}000}{500}$$

$$15{,}000\ \text{E} = 500 \times 3300$$

$$\text{E} = \frac{500 \times 3300}{15{,}000} = 110 \text{ volts}$$

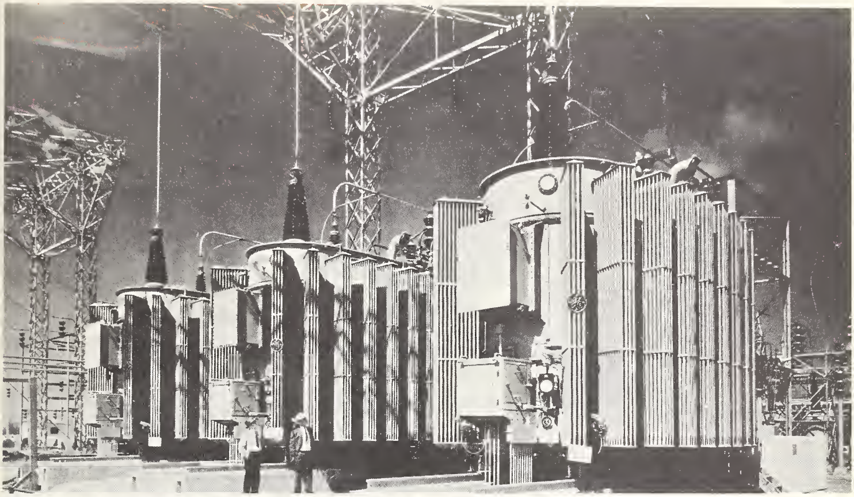


Fig. 28-7. Three large step-down transformers capable of handling 33,000 kilowatts of power. These transformers receive electricity from a high-voltage transmission line, stepping it down from 230,000 volts to 13,800 volts for local distribution.

What is the current in the secondary if the primary coil current is 8 amperes?

$$\frac{\text{Current in secondary (or I)}}{8} = \frac{15{,}000}{500}$$

$$500\ \text{I} = 8 \times 15{,}000$$

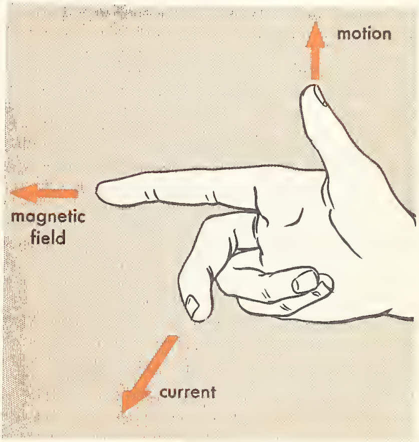
$$\text{I} = \frac{8 \times 15{,}000}{500} = 240 \text{ amp.}$$

The step-down transformer in this example has reduced the voltage from 3300 to 110 volts and increased the current from 8 to 240 amperes.

Direct-Current Generators. How do direct-current (D. C.) generators differ from A. C. generators? We have already seen that in large commercial generators an exciter (small generator) is used to provide current for the rotating field. This exciter must be a direct-current generator so that the field in the rotor of the large generator will always be in the same direction. A D. C. (direct-current) generator is only slightly different than the A. C. (alternating-current) generator. The difference is in the way the induced current from the armature is fed to the brushes. The connection from the ends of the armature coil to the brushes must be reversed every time the current induced in the armature changes direction so that the current, flowing through the brushes, is always in the same direction. A split-ring commutator will do this.

Let us see how the split-ring commu-

Fig. 28-8. The right-hand generator rule.



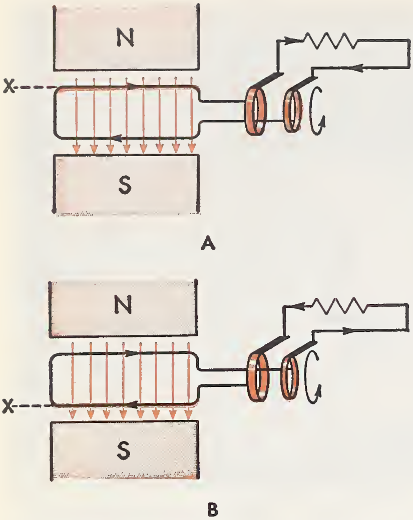


Fig. 28-9. In an A.C. generator the current reverses as it flows through the brushes. Notice the change in position of wire X.

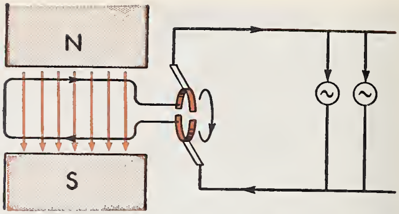


Fig. 28-10. In a D.C. generator the current through the brushes is always in the same direction.

tator works. The direction in which the induced current flows can be determined by ***Fleming's right-hand generator rule.*** (See Fig. 28-8.) Extend your thumb, forefinger, and middle finger of your right hand so that they form right angles with one another. Then, if your thumb points in the direction of the motion of the armature wire, and your forefinger in the direction of the magnetic field, your middle finger will point in the direction of the current flow.

In position (*a*) of Fig. 28-9 note the direction of the current flow in the armature and through the brushes as indicated by the right-hand rule. When the armature reaches the position shown in (*b*), the current induced in it will be in the opposite direction. Check with the right-hand rule as shown. Now the connection of the ends of the armature coil to the brushes is reversed by the action of the two insulated commutator segments rotating with the shaft. In a D.C. generator split-ring commutator segments must be correctly placed with respect to the armature coil, so that they will switch from one brush to the other at the very instant the direction of the current changes in the armature. Thus the output current of the D. C. generator is always in the same direction. In the type described in Fig. 28-10 the current is a ***pulsating*** rather than a steady current. To obtain a steady current in commercial generators many armature coils and many pairs of commutator segments are arranged in succession about the armature.

D. C. generators have many uses. Direct current is necessary in refining metals such as copper, in the manufacture of aluminum and magnesium, in charging storage batteries, in the electrolysis of water, and in electroplating.

Every automobile has a small direct-current generator that provides electricity for the ignition, lights, horn, heater, and radio when the engine is running. The engine turns the armature of the generator which, in turn, supplies the current. The generator also recharges the storage battery which supplies the electricity to start the motor.

Sometimes the shaft of an alternating-current motor is directly connected to the armature shaft of a direct-current generator. In this way alternating-current electricity is used to generate direct current. This device is called a ***motor-generator*** set.

Electric Motors. Michael Faraday, about 140 years ago, discovered that mechanical motion was produced when a wire carrying a current was placed in a magnetic field. (See Fig. 28-11.) This led

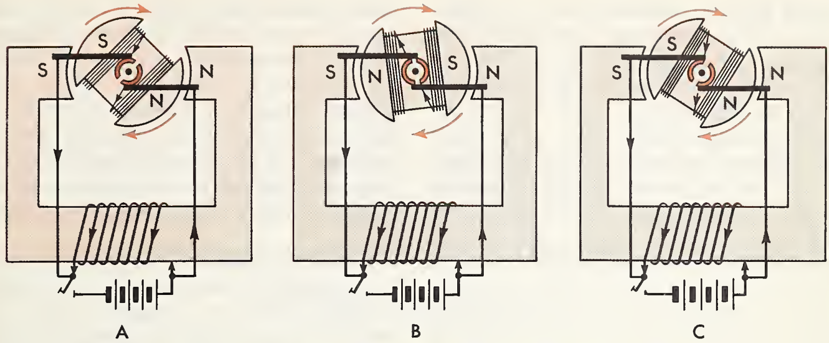


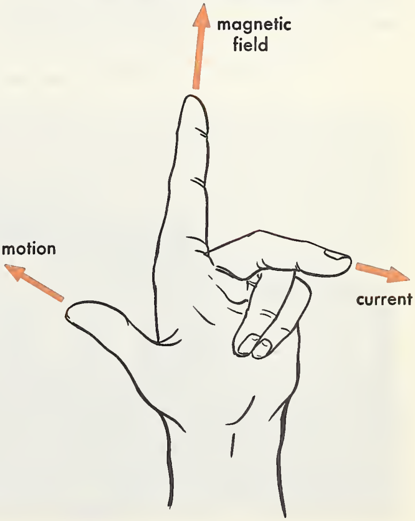
Fig. 28-11. A motor changes electric energy to mechanical energy.

to the invention of the ***electric motor, a device that changes electrical energy into mechanical energy.***

How does an electric motor operate? Motors make use of the interaction of two magnetic fields in converting electrical energy into mechanical energy. The electric motor, like the generator, has an armature and a field. The field may be produced by either permanent magnets or electromagnets. Direct-current motors have commutators and brushes. Current is fed to the rotating armature. Fig. 28-11 shows a simple D. C. motor. The current is fed through two brushes to the slip-ring commutators. When the circuit is complete (switch thrown) the current in the field coil flows as the arrows indicate and the U-shaped field magnet assumes the polarity shown in (*a*). At the same time, current flows through the brushes and through the split-ring commutator segments in such a direction as to give the armature the polarity shown. The two south magnetic poles adjacent to each other will repel, causing the armature to rotate in a clockwise direction until it reaches position (*b*) of Fig. 28-11. At this time each brush is about to change contact from one segment of the split-ring to the other. The two split-ring segments are mounted directly on the armature shaft (but insulated from it and from each other) and so rotate with the shaft. This change causes a reversing of the direction of flow of the current around the armature and so the polarity of the armature changes to that shown in (*c*) of Fig. 28-11.

Fleming's left-hand rule for motors can be used to check the direction in which the armature moves. Extend the thumb, forefinger, and middle finger of the left hand at right angles to each other. Place your forefinger in the direction of the magnetic field (toward the south pole of the field magnet) with your middle finger

Fig. 28-12. The left-hand motor rule



in the direction of the current in the conductor. Your thumb will then indicate the direction of motion of the armature of the motor. Try this with the diagrams as given in Fig. 28-11.

The armature of the motor shown will continue to rotate in a clockwise direction since the polarity of the armature is continuously changed at the proper points during its rotation, every 180 degrees. The type of motor described here is often called a ***universal*** motor, since it can operate on either A. C. or D. C. The starting motor for your auto is an example of a D. C. motor, and the vacuum cleaner motor is an A. C. one.

Notice that D. C. motors and generators are identical. In fact, if a D. C. motor has a permanent magnetic field, it can be used equally well as a generator. In a generator mechanical power is put in and electric power is taken out, instead of electric power going in and mechanical power coming out as in a motor.

Large commercial motors, like D. C. generators, usually have many armature coils and commutator segments to give smoother operation.

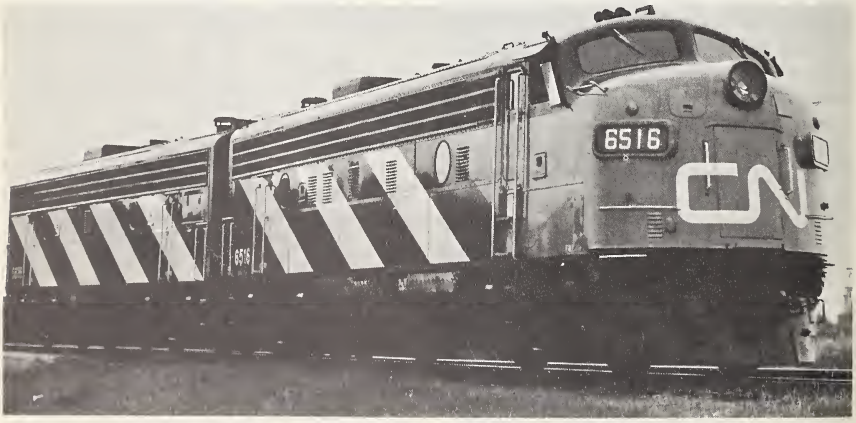
Advantages of Electric Power. Electricity is especially convenient as a source of mechanical power. Steam engines and internal-combustion engines are usually bulky, involve dirty fuel, are noisy and have a great deal of vibration. Also, unpleasant fumes may be given off. Electric motors, on the other hand, are smaller, cleaner to operate, involve no dirty or messy fuel and operate with a minimum of noise and vibration. Only wires are required to carry their invisible electric power, and maintenance is fairly simple.

Electric Locomotives. Large electric motors which get their electric energy from overhead wires are sometimes used instead of steam engines to power locomotives. Fig. 28-13 shows a modern electric locomotive.

The diesel-electric locomotive is one of the smoothest-operating, fastest, and most easily controlled modern locomotives. The crankshaft of the diesel is directly connected to the shaft of an electric generator. The mechanical power from the diesel rotates the armature of the generator, and the electric power thus produced is used to run the electric motors which turn the drive wheels of the locomotive and pull the train.

Electric Power and Energy. Electric motors and generators, like other machines that use energy, are often rated in terms of their power. Electrical devices may be rated in terms of watts or horsepower.

Fig. 28-13. A diesel-electric passenger locomotive consisting of two separate units.



Most motors are rated in horsepower. Sometimes we need to convert horsepower to watts or the other way round. To do this, we should know that 746 watts are equivalent to one horsepower; or that one kilowatt (1000 watts) is equal to 1.34 horsepower. It follows, that if a single horse were used to turn the armature of a generator to provide electric light, he would be able to keep about seven 100-watt lamps burning.

A typical problem involving the cost of electric energy is as follows: *A motor rated at ½ horsepower is operated for 4 hours. How much electrical energy will be consumed and what will it cost at 5¢ per kilowatt-hour?*

$$\text{Power} = \tfrac{1}{2}\text{ h.p.} \times 746 \text{ watts} = 373 \text{ watts}$$

$$= \frac{373 \text{ watts}}{1000} = 0.373 \text{ kilowatts}$$

$$\text{Energy} = 0.373 \text{ k.w.} \times 4 \text{ hours} = 1.492 \text{ k.w.h.}$$

$$\text{Cost} = 1.492 \text{ k.w.h.} \times \$0.05/\text{k.w.h.} = \$0.075$$

Things to Remember

A changing magnetic field about a coil of wire induces a current in the coil. This is called ***electromagnetic induction.***

An ***electric generator*** converts mechanical energy into electrical energy.

A simple ***A. C. generator*** contains a field magnet, an armature, slip rings, and brushes.

A ***D. C. generator*** contains a split-ring commutator.

An ***electric motor*** converts electric energy into mechanical energy.

A simple ***D. C. motor*** consists of a field magnet, an armature, a split-ring commutator, and brushes.

Power in watts is equal to the current multiplied by the voltage.

Motors are usually rated in horsepower; one horsepower equals 746 watts.

Questions

GROUP A

1. What advantages do electric motors have over gasoline engines?

2. Make a list of the various household uses of electric motors.

3. What would be the power output in watts of a generator delivering a current of 40 amperes at a voltage of 6000 volts?

4. Name the important parts of an electric generator.

5. Name the important parts of an electric motor.

6. How much energy is used by a 150-watt light bulb that burns for 5 hours?

7. Describe the construction of a transformer.

8. A step-up transformer with 3000 secondary turns and 100 primary turns has a voltage of 1000 volts across its primary. What voltage is induced in the secondary?

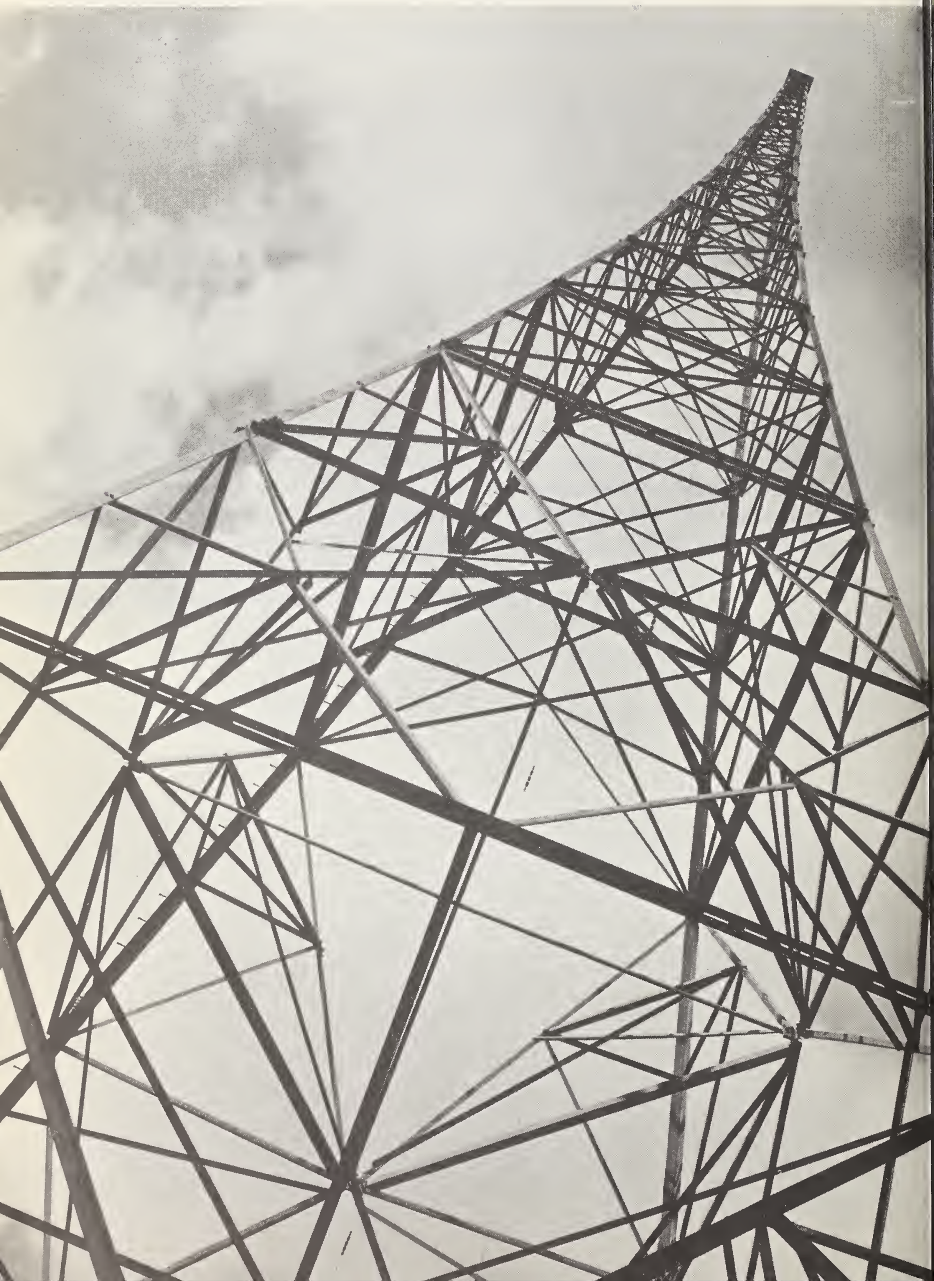
GROUP B

9. How is electricity obtained (*a*) from coal (*b*) from water?
10. Describe how Faraday generated an electric current.
11. An electric motor is rated as 5 horsepower. How many watts of electric power does it use?
12. How many watts are used by an electric motor that lifts a 3000 pound elevator to the top of a 500 ft. building in two and a half minutes?
13. How much does it cost to run a ½ horsepower motor for three hours if electric energy costs 5 cents per kilowatt-hour?
14. Why is the electric current generated in a large alternator taken from the stator rather than from the rotor?
15. A transformer is supplied with 110 volts at 10 amperes on its primary, and delivers a current of 0.5 ampere from its secondary. What voltage is supplied by the secondary?

Things to Do

Visit to electric power plant · Visit an electric power plant in your community. Organize a group and make arrangements for the visit. Plan carefully with your teacher. You will want to look for: (*a*) the coal-feeder system, (*b*) the boilers, (*c*) the condensers, (*d*) the steam turbines, (*e*) the alternators, (*f*) the exciter for producing a magnetic field, (*g*) the switchboard control, and (*h*) the transformer yard. Find out, if you can, the output voltage and current and the voltages and currents at the substations and pole transformers near your home.

Communication



LOOKING AHEAD

Until relatively recently communication by sound was limited to the distance the human voice would carry. At a public meeting, such as a track meet, a man with a booming voice usually made the announcements. Within the past 50 years, however, great advances have been made in sound communication. Even a quiet voice can be projected for hundreds of miles and heard by thousands of people. Moreover sounds can be "stored." That is to say, the human voice can be recorded, and later reproduced, perhaps years after the recordings were made.

All sounds are caused by molecular motion. A vibrating violin string, for example, sets air molecules in motion. The moving air molecules pass through the ear canal and strike against the eardrum causing it to vibrate. This is the way musical sounds are heard. They are transmitted from the vibrating instrument, through the air, and into the ear.

The telegraph, telephone, radio, and television are electrical communicators. In some the transmission medium is a wire, in others it is wireless; in some the electrical signals (which convey information) are changed to audible sound; in others to visible pictures. But whatever the method of communication there are four essential operations common to every system. These are (1) generation, (2) modulation, (3) transmission, and (4) detection. In the telegraph, for example, a battery generates a current, a key modulates the current, a wire transmits the current, and a sounder detects the current. In a telephone there are the same four operations. Modulation of the current, however, is done by a microphone instead of a key, and the current is detected by a receiver. The receiver changes a fluctuating current back into sound.

In radio communication there are neither wires nor cables. Radio is one of the most fantastic scientific achievements of our age. By radio the human voice can encircle the earth in less than a second. It is used to communicate news and for entertainment. But it does far more than this. It plays an important role in navigation by sea and air. Radio instruments can indicate the altitude of a plane or, in fog, they can safely guide a ship or a plane to port.

← *A broadcasting tower. A transmitting antenna is at the top of the tower.* (Ewing Galloway)

29

WHAT IS SOUND?

Have you ever stopped to think that we live in a world of sounds. Without sound we would have difficulty in communicating with one another. Man depends so much on the spoken word that deafness is a great handicap. We hear loud and objectionable noises as well as pleasant harmonious sounds and the voices of our friends. At times we are not conscious of any sound at all, but even when all seems quiet there are sounds that we can just hear but not identify. In absolutely quiet surroundings, such as a sound-proof laboratory (Fig. 29-1), we can even hear our own heartbeat and the faint rushing sound made by blood flowing through our arteries and veins. In fact, the ear is almost sensitive enough to hear the faint sound made by molecules of air as they rush about in random motion, bombarding our eardrums.

How Sound Travels. Sound is a form of energy in motion, and it travels from a source, such as a musical instrument, through matter to our ear. Though all matter conducts sound, some substances are better conductors than others. Steel, for example, is a good conductor. Sponge rubber and fiberboard, on the other hand, are poor conductors, since they absorb a great deal of the sound energy and convert it into heat energy.

How is a substance like steel able to conduct sound energy? The following demonstration will show how sound travels along a steel rod:

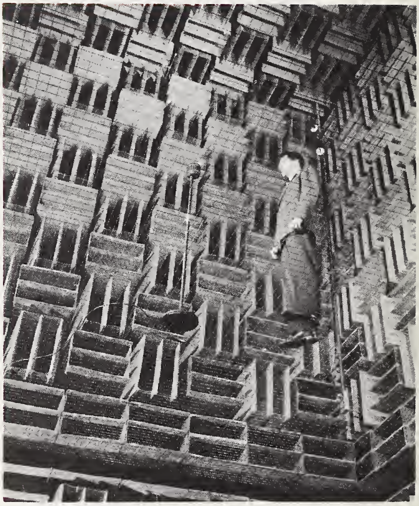


Fig. 29-1. A soundproof room used in sound experiments. To eliminate all surfaces that would reflect sound, the walls, ceiling, and subfloor are lined with special sound-absorbing materials. The unusual floor resembles the hitting surface of a tennis racket and consists of high-strength steel wires strung under high tension between the walls.

Demonstration

29-1. The conduction of sound by steel

Clamp a heavy steel rod in a vise and hang a steel ball from a string so that it just touches one end of the rod as shown in Fig. 29-2. Give the left end of the rod a sharp blow with a hammer. What happens?

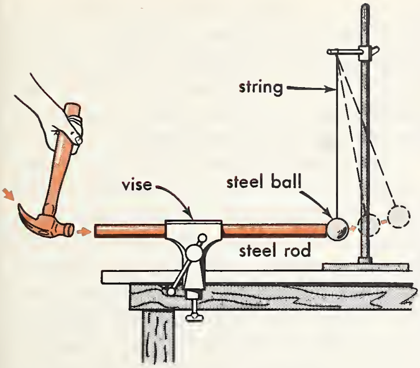


Fig. 29-2. A hammer blow on one end of a steel rod compresses the steel and sends a compression wave down the rod, which pushes the steel ball at the far end.

The ball moves away from the right end and yet the rod, as a whole, is firmly clamped in the vise and has not moved. Energy must have flowed along the bar or the ball would not have moved.

How does this energy flow? When the rod was struck by the hammer, the steel at this end was compressed and the molecules pushed closer together. This compression, or push, was passed along the whole length of the rod from molecule to molecule. As each molecule was, in turn, pushed to the right a tiny distance, the elasticity of the steel pulled it back to its normal position. Thus a wave of motion (or energy) moved rapidly along the rod from molecule to molecule until, at the far end, the molecules in the rod were made to push against the steel ball, so the ball moved rapidly away.

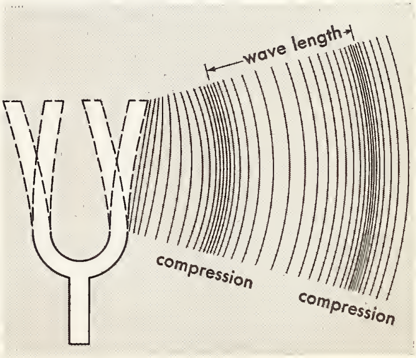
In the conduction of sound energy, one should be aware of the fact that energy does not travel instantaneously down the rod, it takes an appreciable time for each molecule to move to the right and push against its neighbor. However, in steel, sound is conducted so fast (3 miles per second) that the hammer blow and the movement of the ball appear to be simultaneous.

Sound is conducted through air in the same manner as through steel, except that it moves slower through air since the molecules are farther apart and it takes longer for a compression to pass from molecule to molecule. When you clap your hands, the air between them is suddenly compressed and the compression moves out in all directions.

What happens when a tuning fork vibrates? Its prongs push against the air many times a second as they vibrate back and forth. These compressions follow one another through air just like waves in the ocean, only much more rapidly. (Fig. 29-3.) Indeed, every vibrating object sends out sound waves in all directions. When these waves strike the ear drum, they cause vibrations that affect certain nerves that carry impulses to the brain. That is how we hear sound.

Sound Waves. A good idea of the transmission of sound from molecule to molecule is given by the following class experiment. A group of about a dozen students should stand in line, one behind the other, each member placing his hands on the shoulders of the student in front of him. Each student represents a molecule. The student (molecule) at the rear now pushes the shoulders of the student (molecule) in front of him who then in turn pushes the student in front of him, and so on, up the line. It will be noticed

Fig. 29-3. How sound waves are produced by a vibrating tuning fork.



that a single wave, or compression, passes along the lines. Now the experiment is repeated except that the person at the rear should push and then pull at a regular rate; each person up the line passing on first a push and then a pull. Evenly spaced waves will be seen following one another along the line. You will notice that in this model of sound waves, the "molecules" move back and forth along the line of movement of the waves. Such waves are called ***longitudinal waves.*** All sound waves are longitudinal. Ocean waves (and light waves) are different. The water moves up and down at right angles to the direction in which the waves travel. Such waves are called ***transverse waves.***

Notice that in all forms of wave motion, it is energy alone that is carried along by the waves. The matter itself simply vibrates back and forth about a fixed position.

Sound waves obviously cannot travel through a vacuum. Why not? They are transmitted by molecules of matter, and there is no matter in a vacuum. Even sound from the loudest noises produced on earth could never travel beyond the earth's atmosphere. Hence it is impossible for them to be heard on another planet.

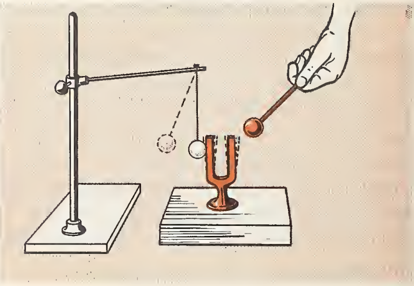
The Speed of Sound. Sound travels through air at about 1100 feet per second or 750 miles per hour, its speed decreasing slightly with a decrease in temperature. It travels faster through solids and liquids than it does through gases. Can you suggest a practical way of measuring the speed of sound in air? Railroad tracks can be used to measure the speed of sound in steel. How would you do it?

The speed of sound is very much slower than that of light, which, you will recall, is 186,000 miles per second. Thus when a ship, seen at a distance of a mile or so, blows its whistle, the puff of steam is visible as the whistle cord is being pulled but the sound is not heard until several seconds later. At track meets the official timers at the finish line of a 220-yard dash must start their stop watches the instant they see the puff of smoke from the starter's gun. They do not wait until they hear the shot. Why not? Because the sound of the shot would not reach their ears until nearly half a second after the race had started.

With nothing more than a watch with a second hand you can determine the distance between you and an approaching thunderstorm. When you see a flash of lightning, measure the time between the flash and the instant you hear the thunder. Using this time and the speed of sound, you can calculate how far you are from the lightning. Suppose, for example, you see a flash of lightning and, ten seconds later, you hear thunder. The sound, traveling at 1100 feet in one second, goes 1100 × 10, = 11,000 feet in the ten seconds. The thunderstorm is therefore about two miles away. Why did we disregard the time for the flash to reach us?

What causes thunder? When a lightning flash occurs, a terrific amount of heat is produced by the very large electric current flowing through the air. This heat, produced almost instantaneously, causes the air along the path of the flash to expand very rapidly sending out a powerful compression wave of sound. Close to, the thunder sounds like a sharp crack but at a distance it rumbles for some time. The rumbling is due to the refraction of the sound by intervening layers of warm and cold air (sound waves

Fig. 29-4. A tuning fork vibrates to and fro when it produces sound, as shown by the motion of the light pith ball.



are refracted just like light waves) and also by reflection from hills and large objects. This reflection and refraction means that some of the sound energy travels by longer paths than the direct route and so arrives at your ear later than the sound that comes to you directly, in a straight line.

Vibration Produces Sound Waves. The following demonstration will show that a tuning fork is vibrating when giving off sound waves although its vibrations are too rapid to be seen.

Demonstration

29-2. A vibrating tuning fork

Support a pith ball by a thread so that it just touches one of the prongs of a tuning fork as shown in Fig. 29-4. Holding the fork firmly at the base, strike the other prong with a small felt-covered wooden hammer. The fork vibrates and produces sound. Watch the pith ball. As long as the fork emits sound the pith ball keeps bouncing back and forth.

The vibrating prong of the fork strikes the pith ball every time the ball comes close enough, making it fly off through the air. When the fork no longer vibrates the pith ball stops moving and no sound is heard, showing that when there is no vibration there is no sound.

Wave Length and Frequency. The distance between any two successive compressions in sound waves is called the ***wave length*** (Fig. 29-5.)

The ***frequency*** of a vibrating source is the number of vibrations (or compressions) produced per second. Thus a tuning fork vibrating 100 times per second will produce 100 compression waves every second, we can also refer to the sound waves themselves as having a frequency of 100 waves or cycles per second.

What is the wave length of the sound produced by the above tuning fork whose frequency is 100 vibrations per second? Since 100 waves are sent out by the fork every second, a new wave must be produced every $\frac{1}{100}$ second. The velocity of the waves is 1100 feet per second so one compression will travel $1100 \times \frac{1}{100} = 11$ feet from the fork before the next compression moves outward from the fork. The wave length (distance from one compression to the next) is therefore 11 feet. Notice that we have calculated the wave length by dividing the velocity of the waves by their frequency. This calculation may be expressed as a very important equation that applies to all kinds of waves:

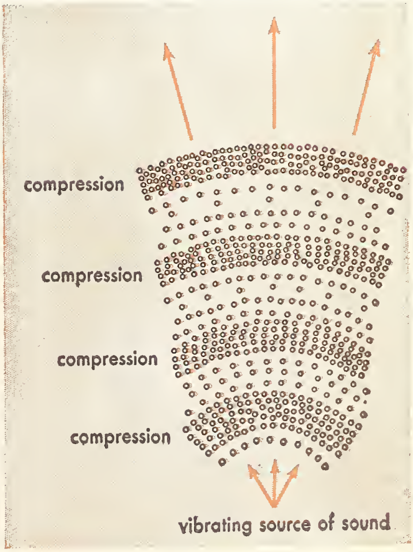
$$\text{wave length} = \frac{\text{velocity}}{\text{frequency}}$$

Or,

$$\text{wave length} \times \text{frequency} = \text{velocity}$$

Pitch. When we say that the notes produced by musical instrument are high or low, we are actually referring to the ***pitch*** of a note. Pitch is determined almost entirely by the number of vibrations

Fig. 29-5. A diagram of a moving sound wave. The compressions and rarefactions of the air molecules move out in all directions from the source at a speed of 1,100 feet per second.



per second produced by the source of sound. In other words, by varying the frequency (vib/sec.), sounds of different pitch are produced. Is a high-pitched sound the result of many or few vibrations per second? The following demonstration will answer this question.

Demonstration

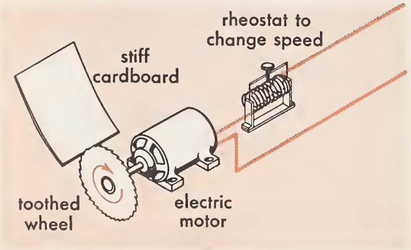
29-3. To produce sounds of different pitch

Fix a toothed wheel to the shaft of a small variable-speed electric motor. Start the motor, and hold a stiff piece of cardboard or sheet metal against the teeth as shown in Fig. 29-6. Change the speed of the motor so that the wheel goes fast and then slow. Notice that as the speed of the wheel changes, the character of the sound also changes.

When the wheel spins rapidly, the card bounces from tooth to tooth many times a second, and a sound of high *pitch* is produced. When the wheel turns slowly, the number of vibrations per second of the card is reduced, and the pitch of the sound produced is low.

Audible Frequencies. The human ear is unable to hear sounds whose frequency is lower than about 20 cycles per second. This lowest audible frequency has a wave length of 1100/20, or 55 feet. Very low frequencies, if they are strong enough, can be ***felt*** by the body but they cannot be ***heard*** as sound. Frequencies higher than 20,000 vibrations per second generally cannot be heard, nor can these high frequencies normally be felt. A frequency of 20,000 cycles per second represents a wave length of 1100/20,000, or .055 feet, or a little more than half an inch.

Fig. 29-6. Producing sounds of different pitch.



Ultrasonics. Certain birds, dogs, and some other animals are able to hear sounds with a frequency higher than 20,000 cycles per second. There is a type of whistle not audible to a human being that can be heard by a dog several blocks away.

Frequencies too high to affect the human ear are called ultrasonic sound. Scientists have discovered that, because of their high energy, ultrasonic waves have many properties that audible sound waves do not have. For example, a piece of cotton, when held in front of a powerful ultrasonic generator, bursts into flame. The generator gives forth no audible sound and has no apparent effect on the human body.

Ultrasonic energy is so intense that the molecular vibration it sets up in liquids is great enough to cause millions of tiny bubbles to form and collapse every second. This process tears dirt and oxides from metal surfaces, providing a new method of cleaning metal surfaces before they are welded or soldered. Ultrasonic energy can make oil and water mix to form an emulsion or homogenize milk, or set dyes or even aid in surgery by breaking delicate tissues. Aluminum can easily be soldered using ultrasonic energy. Every year, the list of useful accomplishments resulting from ultrasonic energy becomes longer.

The Sound Barrier. You learned that one of the most difficult problems that was overcome in designing modern aircraft was to make a plane travel faster than the speed of sound. As a plane approaches the ***speed of sound,*** a very large increase in engine power is needed to increase the speed by just a few miles per hour. At this high speed strange forces appear that make the plane hard to control, and in some cases, even tear it apart. Why should these almost unmanageable

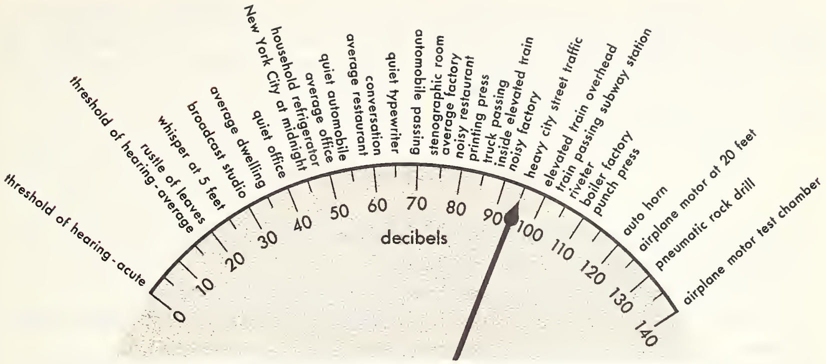


Fig. 29-7. The decibel scale of loudness.

forces appear as the speed of sound is approached?

When a plane is moving through the air it tends to compress the air ahead of it. These compressions are really sound energy and travel out ahead of the plane at the speed of sound. When the plane reaches the speed of sound, however, these compressions are unable to travel out ahead of the plane and so pressure "piles up" in front of the plane. This piling up of pressure is due to a layer of very dense air which must be pushed by the plane. This layer is called a ***shock wave*** and it is this dense compression that causes all the trouble. When the plane flies faster than the speed of sound, it pushes the shock wave ahead of it faster than sound, the wave spreading out to the sides in the shape of a V with the plane at the vertex. This spreading of air waves is just like the bow wave spreading out from the bow of a boat or ship that is moving through the water faster than the speed of water waves on the surface.

This shock wave is very intense. When a supersonic plane flies close to the ground at a speed exceeding sound, its shock wave, spreading out on all sides, is heard as an extremely loud crash and can even break windows at close range.

Loudness and Decibels. The greater the distance through which a sound source vibrates, the more the air molecules are displaced and the louder is the sound. The distance the molecules are displaced is called ***amplitude***. Hence when we wish to describe the loudness or intensity of a sound, we often refer to its amplitude.

An instrument called a ***sound-level meter*** measures the loudness of sound waves in units called ***decibels***. Such instruments are used to investigate noise in factories and large offices and on busy city streets. Continuous exposure to noise can reduce human efficiency and cause fatigue, headaches, and even nausea. A great deal of research has been done to reduce factory and street noises.

Sound Conditioning. Rooms whose walls, ceilings, and floors are constructed of hardsurfaced materials are likely to be noisy, because such materials reflect sound waves and make the smallest sound easily audible. In modern buildings architects use sound-absorbing materials to reduce noise. The sound-reflecting and sound-absorbing properties of the walls, ceiling, and seats of an auditorium determine the ***acoustics*** of the hall. If hard materials are used, there may be too much reverberation, or echo; that is, a sound will travel back and forth over the hall for several seconds before it dies out. This reverberation causes sounds to run together, giving a blurred effect to both voices and music. On the other hand too much absorption by sound-proofing material, makes a hall "dead," and music does not have sufficient brilliance.

Reverberation can easily be demonstrated in a long quiet corridor with hard walls and floors by clapping your hands. The sound can be heard for several seconds before it is completely absorbed.

Things to Remember

Sound travels at a ***speed of 1100 feet per second*** through air.

Sounds are produced by ***vibrating*** objects.

Sound waves are compressions that travel at 1100 feet per second.

The ***wave length*** of a sound wave is ***the distance between two successive compressions.***

The ***frequency*** or ***pitch*** of a sound is the number of vibrations per second.

The ***wave length*** of a wave is equal to its velocity divided by its frequency.

The ***human ear*** is sensitive to frequencies of from 20 to 20,000 vibrations per second.

Longitudinal waves are waves in which the particles of the medium ***vibrate along*** the direction in which the wave travels. Sound waves are longitudinal waves.

Transverse waves are waves in which the medium vibrates at ***right angles*** to the direction in which the wave travels. Water waves and light waves are transverse waves.

The loudness of sound waves is measured in ***decibels.***

Ultrasonic sound waves are sound waves whose frequency is higher than the human ear can detect.

An airplane, rocket or missile produces a ***shock wave*** when it ***breaks the sound barrier*** by traveling faster than the speed of sound.

Reverberation is the re-echoing or multiple reflection of sound waves in a large room.

Questions

GROUP A

1. What is the speed of sound?
2. What is meant by (*a*) the pitch of a sound, (*b*) the loudness of a sound?
3. What is the range of audible frequencies?
4. What is the sound barrier?
5. What are ultrasonics and what use do they have?
6. What causes an echo?
7. How does a vibrating tuning fork produce sound waves?

GROUP B

8. How sensitive is the human ear to sound?
9. A hunter fires a gun and the smoke is seen by a distant observer 3 seconds before the sound is heard. How far is the observer from the hunter?
10. Is the speed of sound ***always*** 1100 feet per second? Explain.

11. A gun is fired 300 yards from a tall cliff. The sound is reflected from the cliff as an echo to the person who fired the gun. How long a time elapses between the firing of the gun and the hearing of the echo?
12. Describe a demonstration that shows that the pitch of a sound depends upon the number of vibrations per second of the source.
13. What is the frequency of a sound wave whose wave length is 2 feet?
14. Water waves with a frequency of 50 per minute have a wave length of 2 feet. What is their velocity?
15. What is the range of wave lengths of sound waves in air that the human ear can detect?

Things to Do

To measure the velocity of sound · The velocity of sound can easily be determined in the country by a small group, under the guidance of a teacher. Borrow a stop watch and a starter's gun from the athletic department. Find a straight stretch of road half a mile or more in length. Measure the distance between two points on the road that are within sight of each other and as far apart as possible. The distance may be measured with an automobile. One group stays at one end with the stop watch, and the other group, with the teacher, goes to the other end of the measured distance. The teacher then fires the gun so that its smoke can be seen by the student operating the stop watch. When the student sees the smoke, he starts the watch, and when he hears the shot he stops the watch.

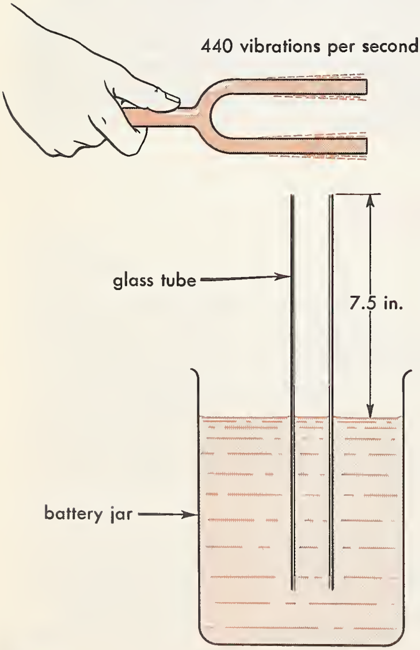
The velocity of the sound can then be found by dividing the distance between the gun and the observer by the time between the sight of the smoke and the sound of the shot, as measured on the stop watch. Make several trials over the given distance and use the average time. The experiment should be done when there is little or no wind. Can you explain why?

30

SOUND INSTRUMENTS

A VIBRATING tuning fork gives forth sound. When the fork is held in the hand, the sound is not very loud. If, however, the base of the fork is pressed firmly against a large hard surface, such as a blackboard, the sound is much louder. Why is this? The fork sets the blackboard in vibration at the same frequency as itself. The vibrations of the blackboard, called ***forced vibrations,*** then reinforce the sound of the fork and make it louder. In the same way the sounding board in a piano strengthens, or reinforces, the tones from the vibrating strings. Without a sounding board the sounds produced by striking the keys would be weak. Sound can be reinforced by any hollow box or chamber, regardless of its shape. The resonant box of a violin, for example, reinforces the sound produced by the vibrating strings.

Fig. 30-1. A tube closed at one end is a resonant column of air one-fourth wave length long.



Resonance. It has been found that a tube closed at one end can reinforce sound, but it reinforces only sounds of particular frequencies. Such a tube is called a ***resonant column.*** The frequency or wave length to which it is resonant depends on its length. The following demonstration shows the relationship between the length of the column and the wave length of the reinforced sound.

Demonstration

30-1. A resonant air column

Almost fill a large battery jar with water and hold a large glass tube in the water. Over the top of the tube hold a vibrating tuning fork whose frequency is 440 cycles per second (the musical pitch of A above middle C, which is the standard for tuning musical instruments). Is the sound of the fork reinforced? Probably not. Now still holding the vibrating fork directly over it, slowly pull the tube out of the water until the sound is strongly reinforced. Now move the tube up and down over a short distance and adjust the height until you have the maximum reinforcement. The air column in the tube is now a resonant length. It measures about 7.5 inches (0.625 ft.).

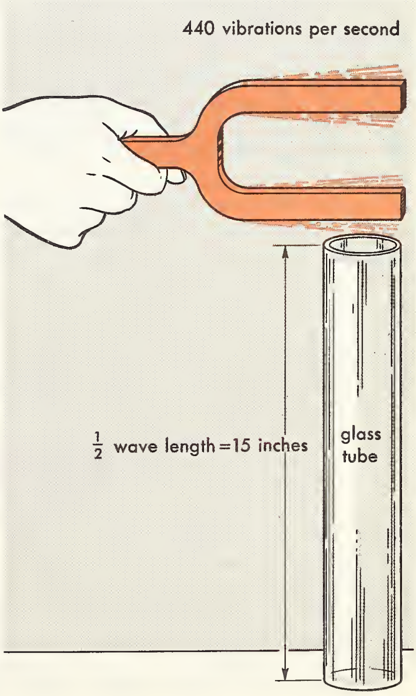
We have already seen that the wave length of a sound can be expressed by the equation $L = \frac{V}{F}$, where L is the wave length, V the velocity of sound, and F the frequency of the wave. By substituting the known values for V and F, we find that in this case $L = \frac{1100}{440} = 2.5$ feet, or 30 inches. The length of the resonant column which reinforces this sound is 7.5 inches, which is one-quarter of 30 inches. We conclude, therefore, that ***the length of a closed resonant column is one-quarter the wave length of the sound it reinforces.***

How can we explain this resonant effect? Resonance depends upon the speed of sound. Consider the vibrating prong of the tuning fork shown in Fig. 48-1. As the prong moves ***down*** it sends a compression down the tube. This compression travels with the speed of sound and so it goes to the bottom of the glass tube, a distance of 0.625 feet in 0.000568 seconds. $\left(\frac{7.5 \text{ in.}}{12 \text{ in./ft.}} = 0.625 \text{ ft.} \quad \frac{0.625 \text{ ft.}}{1100 \text{ ft./sec.}} = 0.000568 \text{ sec.}\right)$

Then it will be reflected, reaching the top of the tube in another 0.000568 sec. That is to say the total time for the compression to go down the tube and be reflected back is 0.00114 sec. In the meantime, the prong which is vibrating 440 times per second takes $\frac{1}{440} =$ 0.00227 sec. for one complete vibration or $\frac{0.00227}{2} = 0.00114$ sec. for one-half a vibration, the same time for the sound to go down the tube and back. If the prong was moving down at the start of our investigation, it will be moving up along with the reflected compression after one-half a vibration or .00114 sec. Thus the sound is reinforced. If the tube had been longer or shorter than 7.5 inches, the reflected compression would not have arrived back at the mouth of the tube in step with the motion of the prong, and the sound would not have been reinforced.

Musical Instruments. A tube that is open at both ends also reinforces sound, but such a tube is a ***half wave length long.*** Both open and closed tubes are used in organ pipes and whistles. The sound is produced in them by a stream of air blown across one end of the air

Fig. 30-2. A tube open at both ends is a resonant column of air a half wave length long.


440 vibrations per second
$\frac{1}{2}$ wave length = 15 inches
glass tube

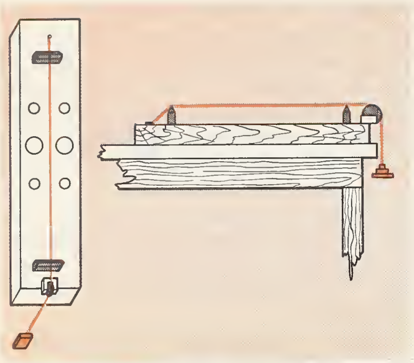


Fig. 30-3. A monochord.

column. This air stream makes the large column of air in the tube vibrate at its resonant frequency. Wind instruments, such as flutes, clarinets, trombones, and tubas, produce sounds in a similar manner.

The pitch of wind instruments is changed by various means, depending on the instrument. A bugle produces sounds of different pitch, depending upon how stiffly the lips are pursed on the mouthpiece and how hard it is blown. These effects combine to make the air column resonate at its fundamental, or at one of its various overtones. Since the frequency of the overtones are only whole number multiples (two times, three times, four times, etc.) of the fundamental, a bugle has only a limited number of widely spaced pitches; it cannot play a scale. A trumpet has valves which cut out short sections of the resonant column, changing its length and thereby giving enough additional pitches to enable a trumpet player to sound all the notes in a scale. In a trombone, the slide, a telescoping section of the resonant column, accomplishes the same thing. Flute, clarinet and oboe players can change the resonant frequency of their instruments, and so increase the number of pitches they can sound by opening and closing holes in the resonant column.

Vibrating strings also produce sound. Violin strings, for example, vibrate when stroked with a bow made of stretched horsehair. ***The shorter the string, the higher the frequency of vibration.*** A string is shortened by pressing it down firmly at different points along its length. A violin is tuned by turning a peg that tightens or loosens each string which also changes its pitch. ***The greater the tension in the string, the higher is its frequency.*** These laws of vibrating strings may be illustrated by means of a monochord as used in the following demonstration:

Demonstration

30-2. The laws of vibrating strings

Make a long, shallow, wooden box. Cut several large holes in the top, and glue triangular wooden wedges near the two ends as shown in the figure. String a wire over the box as shown, with weights hanging from one end. The instrument should be placed on the edge of the table, so that the weights can hang freely.

Observe that the pitch of the instrument rises when more weights are added to the end of the string before it is plucked. Now with a constant weight on the end of the string, notice that a higher pitch is produced when the string is shortened by holding it down with the finger.

The sound produced when a piano is played is also due to vibrating strings. Piano strings are set into vibration by ***percussion,*** that is, by blows from small felt-covered hammers. Drums, cymbals, gongs, and the xylophone are also percussion instruments.

Quality. An organ pipe may produce two sounds of different frequencies at the same time. The sound with the lower frequency is called the ***fundamental.*** The other sound, called a harmonic or ***overtone,*** has twice the frequency of the fundamental. Some sound sources produce many overtones of different frequencies (2, 3, 4, etc. times the fundamental) and intensity. It is the number of overtones and their relative intensity that distinguish the sounds of different musical instruments. The same notes on a clarinet and a trumpet, for example, sound different and are easily distinguishable from each other. This difference in the sound or tone of musical instruments is known as ***quality. The quality of a sound is determined by the nature of its overtones.***

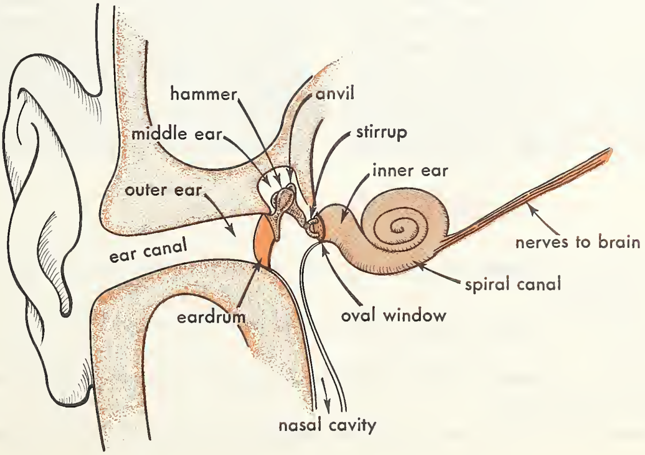
The Voice. The human voice produces sounds by means of vibrating tissue called the ***vocal cords,*** in the larynx, located in the windpipe. These membranes are vibrated by air expelled from the lungs. The tension of the vocal cords is changed by muscular action, producing sounds of different pitch. The mouth and other passages act as resonating chambers. The quality of the voice can be changed by varying the shape and size of these cavities with the aid of the lips, tongue, and teeth.

The Ear. The inside of the ear consists of a series of passages ending in the ***spiral canal,*** which contains thousands of little membranes. Each membrane has a different length, is resonant to a different pitch, and is connected to a nerve ending. When sound waves enter the ear they strike the ***eardrum,*** which starts vibrating. The vibration sets in motion three tiny bones which amplify the sound and conduct it to the spiral canal of the inner ear. There all the tiny membranes resonant to pitches lower than that of the incoming sound vibrate. Their vibrations are transmitted in the form of electrical impulses to the brain, where they are interpreted as sound.

The Phonograph. The phonograph was invented by Thomas Edison. Edison's first phonograph consisted of a cylinder covered with tinfoil against which was pressed a sharp point or needle connected to a metal diaphragm. When the cylinder was rotated, the needle traced a spiral groove around the tinfoil surface. The voice, or other sounds near the diaphragm, made the needle vibrate so that the groove became a wavy line that exactly reproduced the sound variations causing it. This was the "record." To reproduce the original sound the needle was made to retrace the spiral.

The modern phonograph record has tiny wavy grooves that can easily be seen with a powerful magnifying glass in a strong light. The needle vibrations on a modern phonograph disc are converted

Fig. 30-4. A diagram of the human ear.



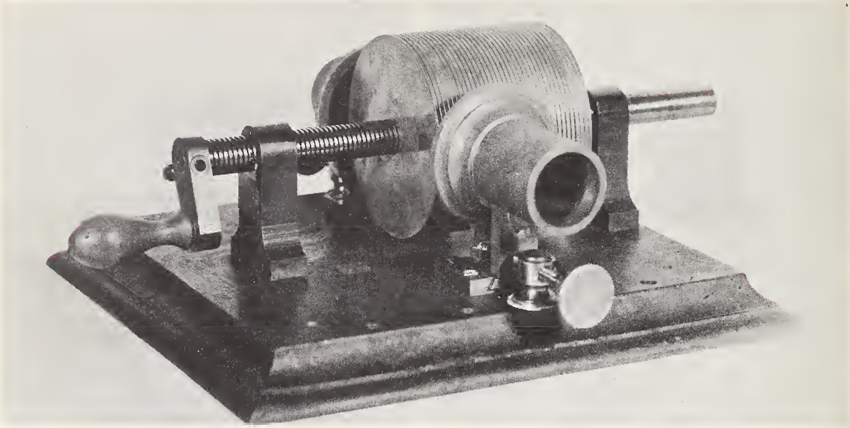


Fig. 30-5. Edison's first phonograph. Sound, entering the tube in front, caused a needle to scratch a wavy spiral track on tinfoil on the cylinder as it was rotated by a crank. When the needle was made to retrace its wavy track, it vibrated, giving out through the tube a replica of the original sound.

into varying electrical currents, which are amplified by vacuum tubes and then reconverted into sound by a loud-speaker.

Tape Recording. Another method extensively used for recording sound is ***tape recording.*** In this system a reel of plastic tape coated with powdered iron is fed past the poles of an electromagnet. The magnetic field of this electromagnet is made to vary by changing the intensity of an electric current in exact correspondence with the sound waves to be recorded. This current is produced by a microphone and vacuum-tube amplifier. The changing magnetic field magnetizes the tape by varying amounts as it moves along, producing a reel with a permanent magnetic record of the recorded sound. When the tape record is "played," it is fed past a coil of wire, inducing a varying current in the coil that is amplified by vacuum tubes. The current is fed to a loud-speaker, thereby producing the sounds originally recorded on the tape.

Although the reels are cumbersome and not as convenient as the more common disc records, tape recording has two advantages: it plays continuously for any length of time, and the tape can easily be run through a demagnetizer that "wipes off" the recorded sounds so that it can be used again. Radio stations often make tape recordings of programs to be broadcast at a later time. Phonograph records are also first recorded on tape and later transferred to plastic records.

Things to Remember

The length of a ***closed*** resonant tube is ***one-quarter*** the wave length of the sound it reinforces.

The length of an ***open*** resonant tube is ***one-half*** the wave length of the sound it reinforces.

The pitch of a ***vibrating string*** is raised when it is shortened and lowered when the tension is decreased.

Overtones occur when more than one frequency is produced at the same time.

Most sound waves are very **complex** because of the presence of overtones.

The **quality,** or characteristic sound, of a given musical instrument depends upon the number and intensity of its overtones.

The **phonograph** was invented by Thomas Edison.

Questions

GROUP A

1. How can sounds be reinforced?
2. A closed tube reinforces a sound wave whose wave length is 4 feet. How long is the tube?
3. How does the ear function?
4. How does the voice produce sounds?
5. What determines the frequency of a vibrating string?

GROUP B

6. How is it possible for a bugle to sound different pitches?
7. To what frequencies will the following columns be resonant: (*a*) a closed tube 5 feet long (*b*) an open tube 2 feet long?
8. Why does a trumpet sounding a certain note sound different from a saxophone playing the same pitch?
9. What is the difference between **forced** vibrations and **resonant** vibrations?
10. What is the length of an open tube that reinforces a sound whose frequency is 200 vibrations per second?
11. What would be the length of closed organ pipes that resonate to the lowest and the highest sounds audible to the human ear?

Things to Do

To measure the speed of sound by resonance · The fact that a closed tube must be a quarter wave length long to reinforce a sound of a given frequency can be used to measure the velocity of sound.

Determine the length of a tube that resonates to a 512-vibration-per-second tuning fork by the method used in Demonstration 30-1. Find the velocity of sound by calculation from the formula *velocity equals frequency times wave length.*

31

THE TELEPHONE AND SOME ELECTRONIC DEVICES

One of the wonders of our modern world is the ease with which men can communicate with one another, even when separated by thousands of miles. A businessman in New York can lift his telephone, give a number and, in a few minutes, talk to a business associate in London, Capetown, Melbourne, or any other part of the world having regular telephone service.

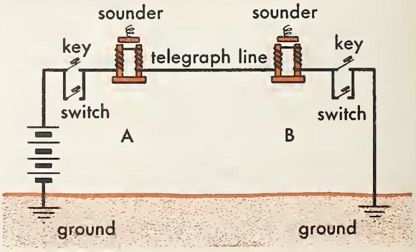
Only a little over a hundred years ago there were no telephones, telegraphs, or radios. As late as 1860 overland communication in much of the United States was still by means such as the famous Pony Express. Then came electricity which made possible the telegraph, telephone, radio and television.

The Telegraph. The telegraph was invented by Samuel F. B. Morse in 1832. Morse first demonstrated his invention by sending over a wire from Baltimore to Washington the now famous message "What hath God wrought!", an appropriate portent to our modern era of rapid communication. Morse's telegraph has changed little in a century, but it has been almost entirely replaced by teletype.

A simple telegraph system includes a ***key*** and ***sounder***. The ***key*** is an electric switch which is normally kept open by a spring and is closed by pushing down on a large button. The key may be kept closed without holding down the button, by means of a switch mounted on the side. The ***sounder*** is a horseshoe electromagnet which, when an electric current passes through it, pulls down a soft iron armature against its poles, making an audible click. When current stops flowing through the magnet, the armature is pulled upward by a spring until it hits a stop and gives out another click.

The circuit of a simple telegraph system is shown in Fig. 31-1. When the

Fig. 31-1. A simple telegraph circuit.



circuit is not in use, the two keys are kept closed by their switches. During this time the current flows from the battery through the key and sounder at station A, then along the wire and through the sounder and key at station *B*, and back through the ground to the battery at A. When the operator at *A* wants to talk to the operator at *B*, he opens the switch on his key. He sends his message by dots and dashes simply by pressing down and releasing his key. Dashes are made by holding the key down for a fraction of a second, and dots by holding it down for a shorter interval. Each letter in the alphabet is represented by a different combination of dots and dashes known as the Morse code. The sequence of clicks made by *A*'s key and sounder is repeated exactly by *B*'s sounder. When A has finished sending, he closes his switch and then operator *B* can send his reply. The code that Morse invented, and which made his telegraph possible, is probably a more remarkable invention than the actual instrument itself.

The Teletype. Most long distance communications today are sent by teletype machines like those shown in Fig. 31-2.

Fig. 31-2. Teletype machines have virtually replaced Morse's original telegraph. This is part of a private teletype system operated by a large air line.

The teletype is operated in much the same way as a typewriter. The message is typed out on a keyboard, producing electrical impulses that operate a similar automatic typewriter at the receiving end. The teletype is more accurate than the telegraph. As the message is printed automatically, no operator is needed to receive it and the Morse code is unnecessary.

The Telephone. The telephone was invented by Alexander Graham Bell in 1876. The basic parts of a telephone system are simple and their operation is easy to understand. The two most important parts are a ***microphone*** and a ***receiver.***

A diagram of a simple telephone ***microphone*** is shown in Fig. 31-3A. It consists of a small metal box about an inch in diameter, that is loosely filled with granules of carbon about the size and shape of grains of sand. Attached to the "cover" of the box is a flexible diaphragm. Sound waves striking this diaphragm make it vibrate, and the vibrations are relayed to the carbon granules. When the carbon particles are compressed, their resistance to electric current is reduced; when the pressure is released, they separate and their resistance is increased. This varying resistance causes a varying electric current in a circuit.

A telephone receiver is a small horseshoe electromagnet with a soft iron diaphragm, a fraction of an inch from the poles. When the varying current from the microphone flows through the coils of the magnet, the diaphragm vibrates and reproduces the sound waves that actuated the microphone. Even complex sound waves that strike the diaphragm of the microphone are accurately reproduced in the receiver of a modern telephone. Fig. 31-3B shows a simple telephone circuit consisting of a microphone, receiver, battery, and connecting wires.

Telephone Repeaters. If a telephone circuit is set up to cover many miles, the resistance of the telephone lines is so

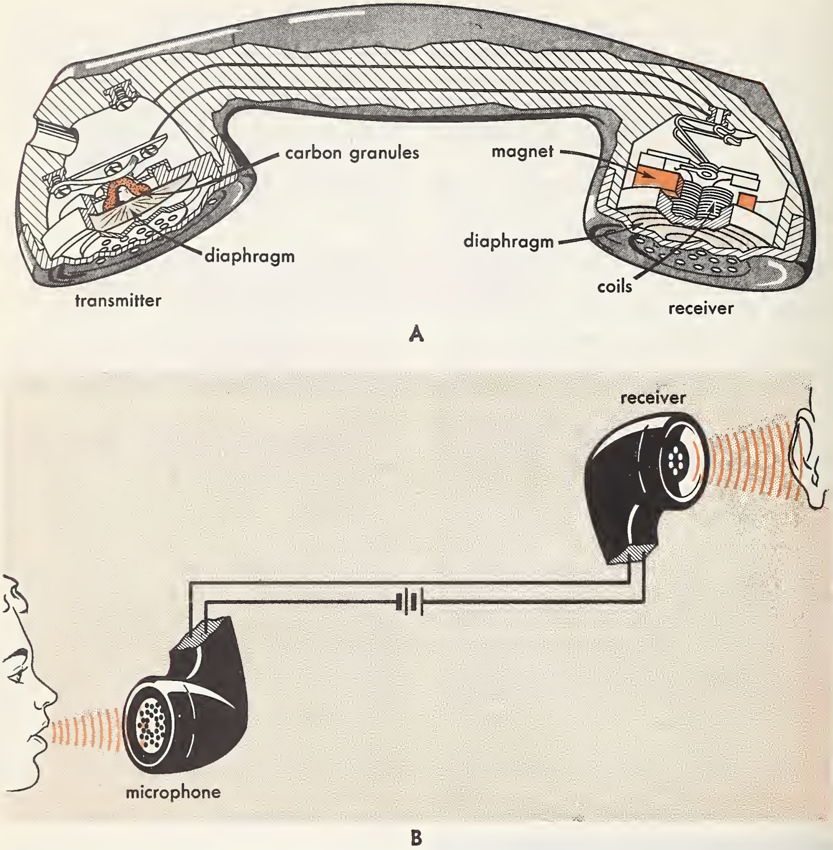


Fig. 31-3. A telephone microphone and receiver (A) and a simple telephone circuit (B).

great that the current flowing in the circuit is too weak to operate the receiver at the far end. This difficulty is overcome by the use of repeaters spaced every 30 miles or so, along the line. By means of vacuum tubes the repeaters amplify or step up the voltage and current, so that there is sufficient power to operate the receiving apparatus at the distant location. In order to understand how vacuum tube amplifiers work, you must first learn something about vacuum tubes themselves.

Vacuum Tubes. A vacuum tube is an evacuated glass or metal tube enclosing several metal elements that control the flow of electrons through the vacuum. Any device that makes use of a vacuum tube is called an electronic device; the study of such devices is called ***electronics.*** The simplest kind of vacuum tube is a diode, which is used to rectify alternating currents, that is, to change them to direct currents.

Diodes. A diode has two important elements, a ***cathode*** and a ***plate.*** Its properties were discovered by Thomas Edison while he was developing his electric light. To see what would happen, Edison sealed an extra wire into a light bulb

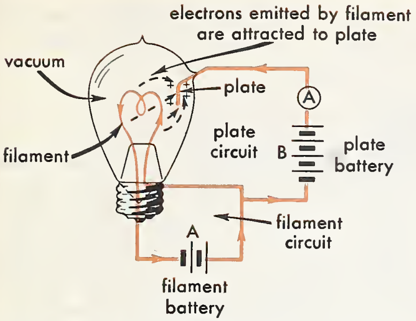


Fig. 31-4. The flow of electrons from the filament in this bulb to the plate wire is called the Edison effect.

in addition to the filament. He then connected a battery and an ammeter between the filament and this extra wire as shown in Fig. 31-4. He found that a current flowed through the ammeter when the filament was lighted. Why was this? If, however, the polarity of the battery was reversed so that the extra wire (a flat plate in modern diodes called simply the ***plate***) was negative, no current flowed through the ammeter. How can this be explained?

This is called the ***Edison effect*** and its explanation is quite simple. When the filament becomes hot, its atoms vibrate so violently that electrons are knocked out of some of the atoms on the surface of the wire and, unimpeded by air molecules, they fly into the evacuated space filling the bulb. The higher the temperature of the filament, the greater is the number of electrons given off per second. Indeed, this process is similar to the evaporation of molecules from a liquid. If the plate has a positive charge, the electrons that leave the filament are attracted to it and flow through the ammeter back to the battery, B. Moreover, as fast as electrons leave the filament or ***cathode,*** as it is commonly called, more electrons from the same battery, B, flow into the filament to take their place. However, if the plate is made negative by reversing the connections of the B battery, it repels electrons from the filament and no current flows. Notice that there are two circuits in Fig. 31-5: the ***plate circuit*** supplied with current by the B battery, and the ***filament circuit*** supplied with current by the A battery. The only function of the filament circuit is to heat the filament so that it will give off electrons. Such diodes (Fig. 31-6) are used in electronic apparatus for changing alternating current to direct current since a diode placed in a circuit acts like a valve, allowing current to flow in only one direction.

Fig. 31-5. In a modern diode the electron-emitting cathode is heated by conduction from a separate filament.

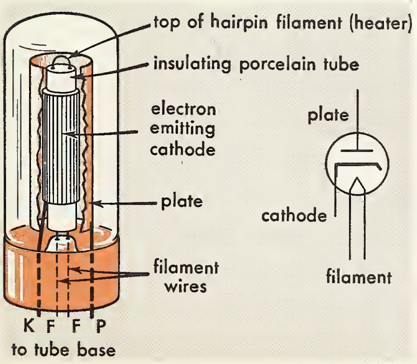
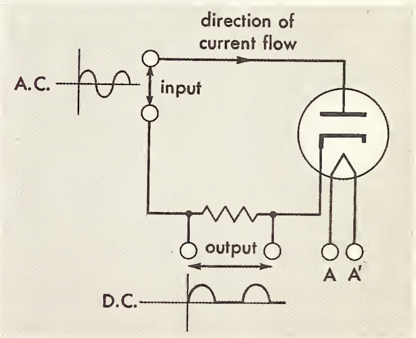

top of hairpin filament (heater)
insulating porcelain tube
electron emitting cathode
plate
filament wires
K F F P
to tube base
plate
cathode
filament

Fig. 31-6. A diode rectifier circuit. When an A.C. voltage is applied at the input, current flows in only one direction through the diode and a pulsating D.C. voltage appears across the output.



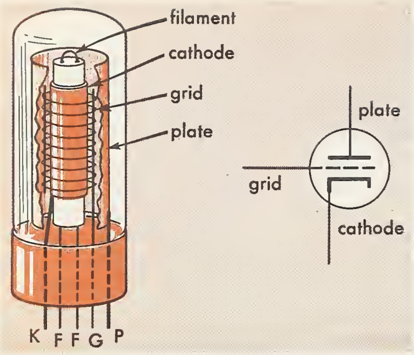


Fig. 31-7. The construction of a triode (left) and its schematic symbol (right).

Triodes. A triode tube, as its name suggests, has three important elements, and is used to ***amplify*** a current. Like a diode, it has a cathode and a plate; in addition it has a ***grid*** of fine parallel wires placed between the cathode and the plate. Lee DeForest was the first to make such a tube. He found that when the grid is charged negatively, the flow of current between cathode and plate is reduced. The electrons moving from the cathode toward the plate are repelled back toward the cathode by the negatively charged grid. As a result, only a few electrons get between the grid wires and reach the plate. If the grid is made positive, however, electrons are helped on their journey to the plate, and the current through the tube is increased.

Demonstration

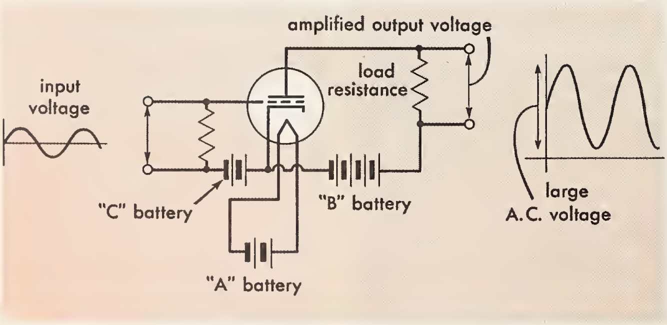
31-1. The construction of a triode tube

Obtain a burned-out triode tube with a glass envelope from a radio service shop. Wrap it in cloth, and break the glass envelope by squeezing it in a vise. With a pair of wire cutters, remove the plate and examine the structure of the tube. Look carefully for the three elements—grid, cathode, and plate—and notice how they are connected to prongs on the base of the tube.

The grid, as DeForest discovered, can be used to control the current and voltage in the plate circuit and, most important of all, ***a small change in the voltage applied to the grid of a tube will produce a large change in voltage in the plate circuit of the tube.*** It is this property of a triode that gives us our modern electronic amplifiers that change small voltages to large, supplying useful voltages in telephone repeaters, phonographs, radios, sound-movie projectors, and many other electronic devices.

A Vacuum-Tube Amplifier. Fig. 31-8 shows the circuit of a simple triode amplifier. The A battery (usually 1 to 6 volts) is needed to heat the filament which in turn heats the cathode of the

Fig. 31-8. The circuit of a simple triode amplifier.



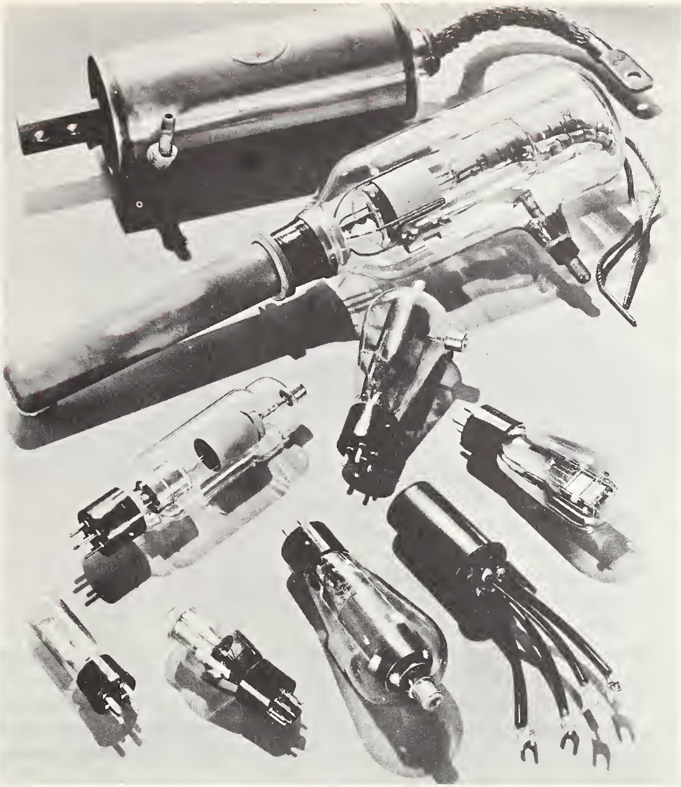


Fig. 31-9. Vacuum tubes used in industrial electronic circuits.

tube so that it gives off electrons. The B battery is so connected that it makes the plate positive and a steady D. C. current flows around the ***plate circuit,*** from the battery through the load resistance, through the tube and back to the battery. You will recall that it is ***electrons*** that actually move, and they flow in the opposite direction from what we call the current. They flow from the battery through the tube, then through the load resistance and back to the battery.

Since the C or bias battery is connected so as to make the grid negative, the electron flow through the tube in the plate circuit is limited. No current will flow through the tube grid circuit (from the C battery through the tube from cathode to grid and then through the grid resistance back to the C battery) because the negative grid repels electrons.

Now suppose a small alternating input voltage (such as might be produced by a microphone or a phonograph pick-up) is applied across the grid resistance. This will make the voltage on the grid of the tube vary, thus causing a varying plate current through the tube. This rather large variation in current flows through the load resistance, producing a varying voltage which is much larger than the tiny A. C. input voltage. Thus we get a larger A. C. voltage out of the amplifier than we put into it. If further amplification is desired, the output voltage from this amplifier tube can be fed into the input of another amplifier tube and increased still further. Electronic amplifiers can be made using several tubes—which amplify very small voltages as many as a million times!

You may have asked yourself, Why go to all the trouble of using complicated and expensive electronic amplifiers when a single transformer can be used to step up voltages? The answer is that a transformer can be designed to step up voltage by large amounts when only ***one*** frequency of alternating current is to be amplified. It is impossible to design a single transformer that will have a large step-up ratio for A. C. voltages of widely differing frequencies. If we wish to amplify the tiny voltages from a microphone with high fidelity we must be able to amplify, equally, a wide range of A. C. frequencies (the range of the human ear is 20 to 20,000 cycles per second). This job can be done conveniently only by electronic amplifiers using vacuum tubes or transistors.

Fig. 31-10. The tiny transistor (right) can do the same job as the vacuum tube (left) and requires much less space and power.

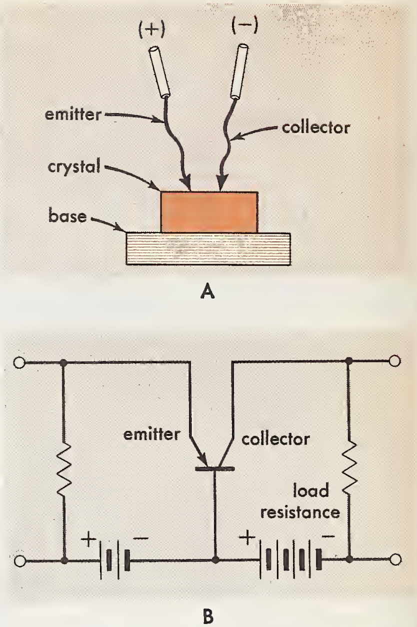


Fig. 31-11. The construction of a transistor (A) and the circuit of a transistor amplifier (B). The signal to be amplified is applied to the emitter circuit at the left, and the amplified signal appears at the right in the collector circuit.

Transistors. Transistors are devices that can be used as amplifiers in place of vacuum tubes. They were invented at the end of World War II by three scientists at the Bell Telephone Laboratories, who were awarded a Nobel prize for their discovery.

A transistor is very small (about the size of a pencil eraser) and requires no filament. It is simply a very thin slice of a ***semiconductor,*** a material that is neither a good electric insulator nor conductor. A silicon crystal containing very small amounts of certain impurities is sometimes used. There are three connections to the transistor corresponding to the cathode, grid, and plate of a triode. The crystal is mounted in a base and this is one connection, the "cathode." In one form of transistors (Fig. 31-11A), two fine wires make contact with the crystal. One is called the ***emitter*** (corresponding to the grid) and the other is called the ***collector*** (the plate). When the voltage on the emitter is changed, the current

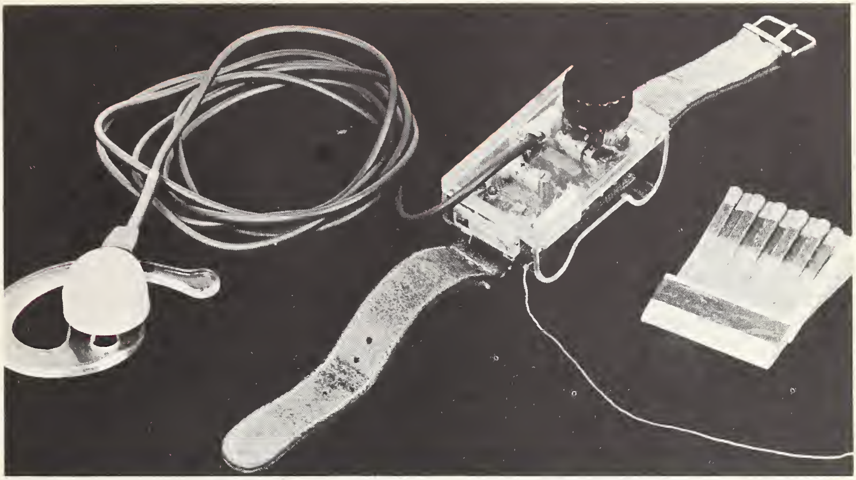


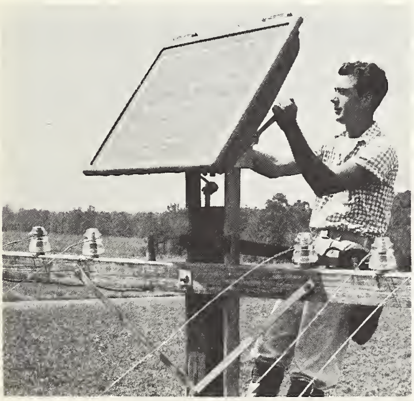
Fig. 31-12. A wrist watch radio using transistors and a tiny battery. The earphone is on the end of the cord at the left.

flowing to the collector is also changed, thus producing a varying voltage across the load resistor.

Transistors are an improvement over tubes because of their very small size. Moreover, as no power is needed to heat a filament, their power consumption is very low and little heat is produced. They are rapidly coming into use in all sorts of electronic apparatus and will soon replace tubes in most devices.

Electricity from the Sun. It has recently been discovered that when light falls on silicon crystals of the type used in transistors, electrons are liberated from atoms in the crystal and can move through the crystal as an electric current. In other words, light energy is converted into electric energy in this device. A solar battery giving useful voltages and power from the sun's radiation can be constructed by connecting a large number of individual silicon cells in series as shown in Fig. 31-13. Such a battery exposed to direct sunlight will give about 100 watts for every square yard exposed to the sun. The efficiency of a ***solar battery*** is rather low; only 11 per cent of the radiation energy falling on the battery is converted into electric energy. Nevertheless there is so much available energy from the sun's radiation, that solar batteries can be very useful despite their losses. Solar batteries are now being placed on telephone poles in remote rural areas. The electric energy produced when the sun is shining is used to charge a storage battery so that electricity will be available

Fig. 31-13. A solar battery being installed to convert the sun's light energy into electrical energy for use in a rural telephone circuit.



24 hours a day to operate a telephone line amplifier. They can also be used in artificial satellites to provide power for a radio transmitter.

Long-Distance Communication. By the use of electronic amplifiers it is possible to talk into a telephone in New York and be heard in San Francisco as clearly as though the telephone line were only a few feet in length. On some transcontinental telephone lines, there are as many as 450 separate repeater amplifiers giving a total amplification of 10^{1000} times (ten followed by 1000 zeros). But let us return to the businessman in New York who wishes to talk to someone in London. How can we amplify signals every 30 miles along a submarine cable lying on the floor of the Atlantic Ocean? Though not easy, it can be done, and such a cable with underwater amplifiers has recently been installed under the Atlantic. Such cables are very expensive to make and they can handle only a limited number of conversations at the same time. Most transoceanic telephoning is accomplished by means of radio links over the ocean. In the next chapter you will learn about radio communication.

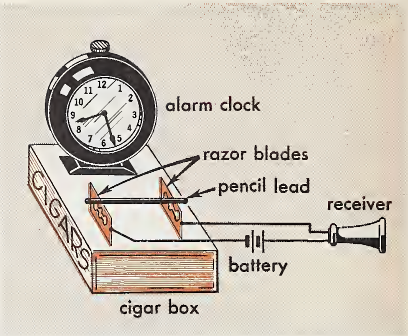


Fig. 31-14. A simple microphone that will pick up the ticks of an alarm clock can be made from a cigar box, two razor blades, and a pencil lead.

Things to Remember

The ***telegraph*** was invented by Morse in 1832.

The ***telephone*** was invented by Bell in 1876.

The flow of electrons from a hot filament, through a vacuum to a plate is called the ***Edison effect.***

A ***diode*** is a vacuum tube that contains a filament, cathode, and plate. Diodes are used as ***rectifiers*** to change alternating current to direct current.

A ***triode*** contains a filament, cathode, grid and plate. A triode can be used to ***amplify*** weak electric currents.

Transistors perform the same function as vacuum tubes but are much smaller and consume much less power.

A ***solar battery*** is a battery of cells that converts energy of sunlight into electric energy.

Questions

GROUP A

1. What are the essential parts of a telegraph?
2. How is a telegraph sounder constructed?
3. Draw the circuit of a simple, two station telegraph circuit and explain how it is used to send messages back and forth.
4. What are the essential parts of a simple telephone circuit?
5. What are telephone repeaters and where and why are they used?

6. What is a diode tube?
7. What is a triode and for what can it be used?
8. What is a transistor?
9. What is a solar battery?

GROUP B

10. Describe the construction and operation of a microphone.
11. Describe the construction and operation of a telephone receiver.
12. Draw the circuit of a simple, one way telephone system and describe its operation.
13. Why will no current flow through a diode when the plate is charged negatively and the cathode charged positively?
14. What is the function of the grid in a triode?
15. Why are triodes used instead of transformers to amplify weak voltages and currents?
16. Draw the circuit of a triode amplifier and explain its operation.
17. What are the advantages of transistors over vacuum tubes?
18. Draw the circuit of a diode rectifier circuit and explain how it changes alternating to direct current.
19. Why is the emission of electrons from a hot filament like the evaporation of a liquid?

Things to Do

To make a simple microphone · A simple, home-made microphone is shown in Fig. 31-14. The base is a cigar box in which two double-edged razor blades are imbedded. An electrical connection is made by soldering a wire to each blade and laying a pencil lead across the edges of the blades. The microphone should be connected in series with a telephone receiver and a 3-volt battery of two dry cells connected in series. If an alarm clock is placed on the box, you will be able to hear it tick by listening in the receiver. Can you explain how this microphone works?

To study the history of communications. Many interesting stories are connected with the development of communications. What can you learn about the invention of the telephone and the laying of the first transatlantic telegraph cable?

32

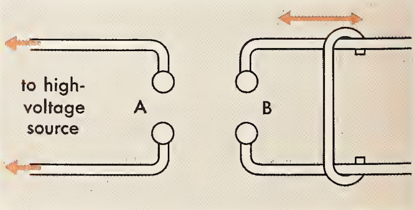
RADIO AND TELEVISION

Radio is one of man's most spectacular inventions. The complex sounds of a symphony can be sent through space to be reproduced at distant places by radio receivers. Radio waves make the transmission possible. These waves, though electromagnetic waves like light, cannot be seen—their wave lengths are much too long to affect the retina of the eye. They travel almost a million times as fast as sound—at the speed of light—and so enable a radio listener in Victoria to hear the voice of a singer in a large auditorium in Saint John a little sooner than the people in the back rows of the hall. Similarly, television enables us, by means of radio waves, to see at a distance far beyond the range of the human eye. How has man learned to put electromagnetic waves to such amazing uses? To answer this question let us go back a century to the time of James Clerk Maxwell, a Scottish mathematician and physicist.

The Discovery of Radio Waves. From his study of light waves, Maxwell predicted that radio waves must exist and have properties similar to light waves. How were his predictions verified? Heinrich Hertz, a German physicist, was the first to demonstrate in the laboratory that such waves could be produced and detected. The apparatus he used to generate and detect these waves was very simple, consisting mainly of two spark gaps (see Fig. 32-1). The spark gap on the left was supplied with a high voltage so that a series of sparks jumped across the gap. The adjustable wire loop and gap on the right was placed a few feet away and, when the sliding connection on the loop was placed at the proper point, sparks would also jump across this gap, though there was no source of power in this "receiver." The energy of the spark had traveled by electromagnetic waves through space from the "transmitter" to the "receiver."

Hertz investigated the properties of these waves and found, as Maxwell had predicted, that they could be reflected by metal surfaces, focused into beams,

Fig. 32-1. Hertz's apparatus for generating and receiving radio waves.



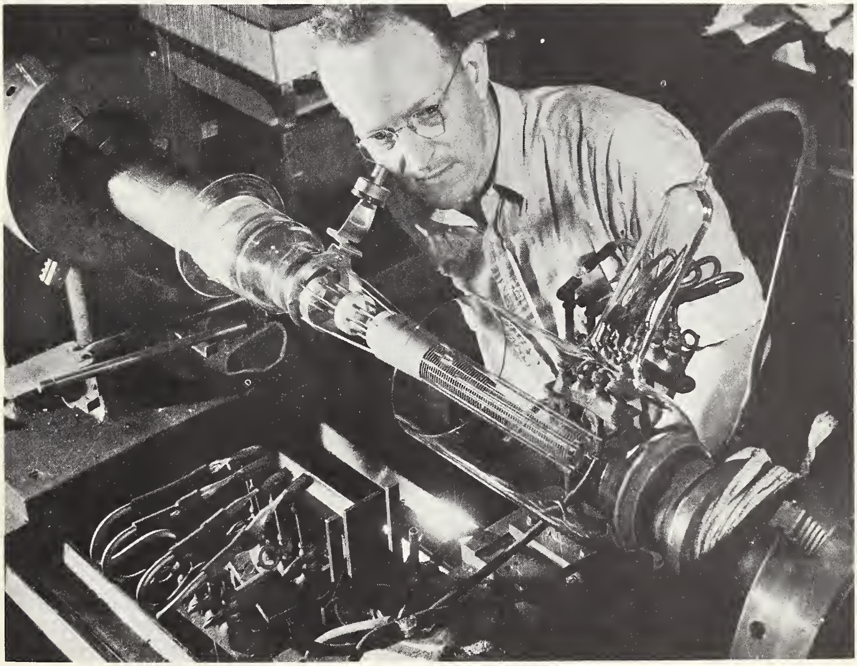
and refracted. However, he never thought of the possibility of using these waves for communication purposes. It was Guglielmo Marconi, an Italian, who first put them to practical use in the form of a wireless telegraph. How did he do this?

Early Wireless Telegraphy. Marconi increased the distance over which radio waves could be sent, and used them for communication. He started in 1893 by sending a message over a distance of 5000 feet. In 1899 he communicated, by ***wireless*** telegraphy, as radio was then called, across the English Channel. In 1901 he succeeded in receiving at St. John's, Newfoundland, signals from a transmitter in England. This event marked the coming-of-age of radio. Soon ships were using Marconi's apparatus to keep in touch with their home ports and to appeal for help in emergencies. One of the first dramatic uses of wireless telegraphy resulted in the arrest of a man wanted for murder. The criminal, fleeing from England to Canada, was arrested as he stepped off the boat at Halifax.

Radio communication did not progress beyond simple wireless telegraphy for some time—not until the triode vacuum tube was perfected. In 1920 station KDKA in Pittsburgh broadcast the first radio entertainment program of voice and music.

Generation of Radio Waves. To understand how an electric spark produces radio waves we must first realize that a spark discharge consists, not of a simple flow of electric charges in one direction, but rather of a rapid vibration of charges across the spark gap. In other words, it is an alternating current flowing across the gap. The frequencies of such alter-

Fig. 32-2. Large vacuum tubes used in powerful radio transmitters are virtually made by hand. Here a glass worker is turning such a tube on a lathe, inserting the cathode before sealing off the tube.



nating currents range from a few thousand cycles to several billion cycles per second. Flowing in the wires of a transmitting antenna, these ***high-frequency currents*** send out electromagnetic waves in all directions through space with the speed of light.

Radio transmitters no longer use electric sparks to generate radio waves. Soon after the discovery of the Edison effect it was found that a triode vacuum tube would generate high-frequency currents much more efficiently. Modern radio transmitters contain many vacuum tubes connected in a complex circuit. A transmitter always contains an ***oscillator tube*** that generates alternating currents of specific frequency, but usually of very low power. These alternating currents are then amplified many thousands of times by a series of amplifier tubes and fed into the transmitting antenna, which radiates this energy over great distances as powerful radio waves. The arrangement of oscillator, amplifiers, and antenna is shown in Fig. 32-3. How can such apparatus be used for communication?

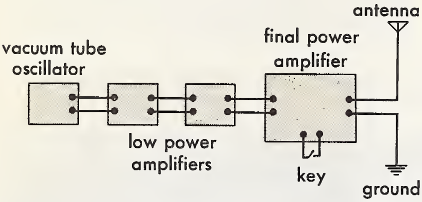


Fig. 32-3. A block diagram of a radiotelegraph transmitter.

A radio transmitter of this kind can be used only for radio telegraphy. The source of high-voltage power to the final amplifier tubes is turned on and off by means of a telegraph key, thus producing a radio wave interrupted according to the dots and dashes of the International Morse code. In order to transmit voice and music, additional equipment is needed.

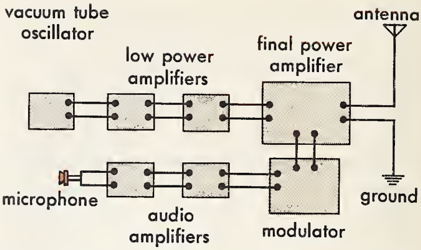
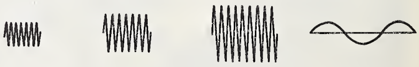


Fig. 32-4. A block diagram of a radiotelephone transmitter.

Broadcasting Voice and Music (Radiotelephony). To transmit voice and music the high-frequency waves must be made to vary in intensity according to the variation in the sound waves of the voice or music. How is this done? First the sound variations are converted into a varying electric current by means of a microphone, just as in a telephone. This varying voice current, or ***audio-frequency current,*** is then increased in power by a vacuum-tube amplifier called a ***modulator.*** This amplified current controls the power applied to the final amplifier of a radio transmitter such as that shown in Fig. 32-4. In this way the intensity of the radio waves is varied in accordance with the sound variations of the voice or music entering the microphone. This type of ***modulation*** (varying of the radio waves

Fig. 32-5. Wave forms in different stages of a plate-modulated radiotelephone transmitter.

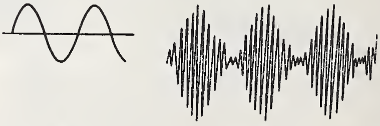


(a) oscillator current

(b) first low power amplifier current

(c) second low power amplifier current

(d) microphone current



(e) modulator current

(f) radio frequency current of final amplifier varied at audio rate by modulator current

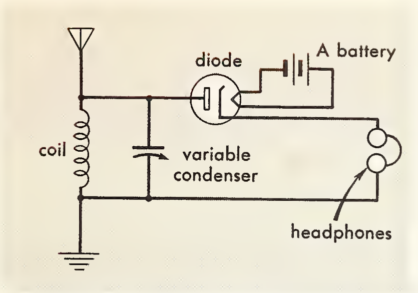
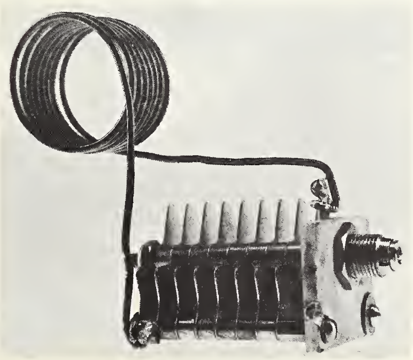


Fig. 32-6. A simple radio-receiver circuit.

in accordance with the sound to be broadcast) is called ***amplitude modulation*** because the amplitude, or intensity, of the waves is varied.

Let us now summarize, with the aid of Fig. 32-5, the operation of a modern broadcast transmitter. Radio-frequency currents are generated by the oscillator tube. (These currents are shown in part *a* of Fig. 32-5.) They are amplified by low-power amplifiers (see *b* and *c*) and still more by the final power amplifier. At the same time the microphone current (see *d* of Fig. 32-5) is amplified by the modulator (*e*) and then used to vary the intensity of the radio-frequency currents in the final amplifier, as shown in (*f*). Thus we have a high-frequency current that varies in intensity being fed into the antenna from which a corresponding radio wave is radiated.

Fig. 32-7. A tuned circuit consisting of a coil and a variable condenser. This circuit is used in a radio receiver to select the desired radio station.



A Simple Radio Receiver. How can radio waves be picked up, detected, and converted into audible sound by a radio receiver? Four parts are needed in a simple receiver: (1) an antenna, (2) a tuned circuit, (3) a detector, and (4) a reproducer. The antenna need be only a wire, between 30 and 100 feet long, insulated from surrounding objects and placed high above the ground. One end of it runs to the tuned circuit, which consists of a coil and a ***condenser*** (see Fig. 32-7). The tuned circuit is connected to the diode detector tube which, in turn, is connected to the reproducer. The reproducer consists of a pair of headphones and is similar to a telephone receiver. Fig. 32-6 shows how this apparatus is connected. An alternating current is induced in the antenna when it intercepts electromagnetic radio waves. This current corresponds exactly in frequency and amplitude variation to the radio waves. In fact it is exactly the same (though much weaker) as the current flowing in the antenna of the radio transmitter that originally produced the radio waves. This induced current flowing through the coil produces a voltage, which causes current to flow through the diode detector tube and the headphones.

The Detector. What is the purpose of the detector? How does it work? First, let us see what would happen if the changing-intensity, high-frequency alternating current passed directly through the headphones. The nature of this current is shown in Fig. 32-8A. The narrow spikes of different height represent the very rapidly changing radio-frequency current. The effect of this current on the diaphragm in a headphone is to push it back and forth at such a rapid rate that the inertia of the metal diaphragm prevents it from moving more than a microscopic distance before the current has changed direction and the diaphragm moves back. The net result is that the diaphragm's

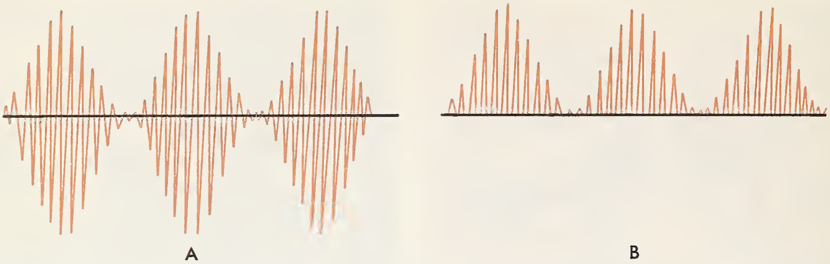


Fig. 32-8. Wave forms of a modulated radio frequency current before (A) and after (B) detection.

motion is too small to produce audible sound waves.

However, the diode allows current to flow in only one direction through the headphones. This kind of current, called a ***rectified*** current is shown in Figure 32-8B. It is really a direct current of rapidly varying intensity. What is the resulting motion of the diaphragm under the action of this current? These very rapid pulses of current always push the diaphragm in ***one direction*** so it now moves back and forth at a relatively slow rate corresponding to the audio change in amplitude of the radio-frequency current. This audio change is slow enough so that the inertia of the diaphragm interferes little with the motion, and sound waves are produced by the diaphragm exactly corresponding to the sound modulation impressed on the original electromagnetic radio wave.

The Tuned Circuit. What is the purpose of the tuned circuit? The tuned circuit selects the particular radio waves to be received. If there were no tuned circuit, all radio stations emitting waves intercepted by the antenna would be heard as a jumble of sound. The coil and condenser combination of the tuned circuit allows only radio-frequency currents of one frequency to build up large enough voltages to operate the receiver. The particular frequency the tuned circuit is resonant to depends upon the size or ***capacity*** of the condenser and the number of turns of wire on the coil. Thus to "tune" a radio receiver (select a particular station) the capacity of the condenser is varied by a tuning knob. The tuning circuit is called a resonant circuit because, like a resonant air column which reinforces one sound frequency, it reinforces one radio frequency. Different transmitters generate their radio waves at different frequencies, so that each can be selected in turn by tuning a radio receiver.

Modern Radio Receivers. The simple receiver we have just discussed could have its sensitivity, or ability to pick up weak signals, increased by using vacuum-tube amplifiers. ***Audio amplifiers*** can be added to increase the power of the signal from the detector. ***Radio-frequency*** amplifiers can be placed between the antenna and the detector to amplify the radio-frequency currents from the antenna before they are detected. Each of the radio-frequency amplifier tubes include a tuned circuit to increase the ***selectivity*** of the receiver. The selectivity is the ability of the receiver to select or receive a weak station without interference from a stronger station operating at nearly the same radio frequency.

A receiver using both audio and radio-frequency amplifiers is usually called a ***tuned radio-frequency receiver.*** Most modern receivers are essentially of this type but include an additional tube, a frequency converter, in what is called a ***superheterodyne*** circuit.

Transmission of Radio Waves over Long Distances. With light waves we can see only as far as the horizon. One might well ask why radio waves, which have proper-

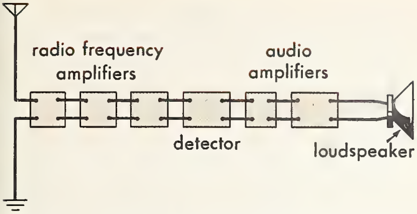
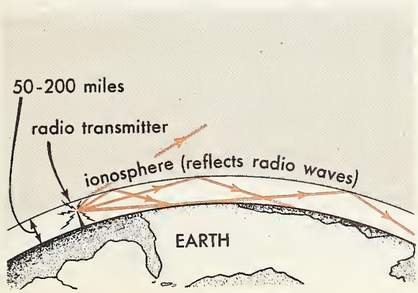


Fig. 32-9. A tuned radio-frequency receiver. Each radio-frequency amplifier stage contains a tuned circuit.

ties similar to light waves, can bend around the curvature of the earth and travel all over the world. They can do this because far above the stratosphere there is a layer of air at very low pressure called the ***ionosphere*** (see Fig. 14-7 in Chapter 14). The ionosphere conducts electricity; and though it cannot reflect light, it does reflect the longer electromagnetic waves of radio around the curved surface of the earth as shown in Fig. 32-10. Actually, not much radio energy is reflected by the ionosphere from stations in the standard-broadcast band. Most of the energy from these stations reaches us directly through the atmosphere and the earth's surface. The reflection of the ionosphere is most efficient in the short-wave band, and stations operating in this range are used for transoceanic telephone service.

Since there are many thousands of radio transmitters now operating all over the world, the lower frequencies are overcrowded. Therefore such new developments as FM and television (TV) must operate on the very high frequencies where there is more room. Since the range of transmitters operating on these frequencies is only as far as the natural horizon (such high-frequency waves not being reflected by the ionosphere), FM and television transmitting antennas are usually placed as high as possible, on the tops of tall buildings and hills, to extend their range. For the same reason, FM and television receiving antennas should also be placed as high above the ground as possible.

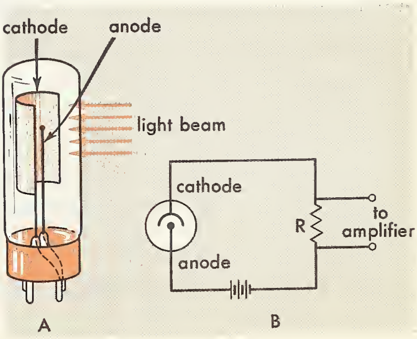
Fig. 32-10. Radio waves are reflected around the curve of the earth by the ionosphere.



Television. Television is the transmission of light images by means of an electric current and radio waves to a distant point. Until some device could be found that would change variations in light intensity into electric current variations, television could not be developed. Not many years ago such a device was invented. It is called a photoelectric cell and it is an essential part of all television systems.

The Photoelectric Cell and Sound Movies. A photoelectric cell is a vacuum tube containing a cathode and anode. The cathode, unlike most vacuum tubes, is not heated and normally does not give off any electrons. However it is coated, with a substance such as cesium, which gives off electrons when exposed to light (Fig. 32-11). The light energy absorbed

Fig. 32-11. The construction of a photo-electric cell (A) and a simple photo-electric cell circuit (B).



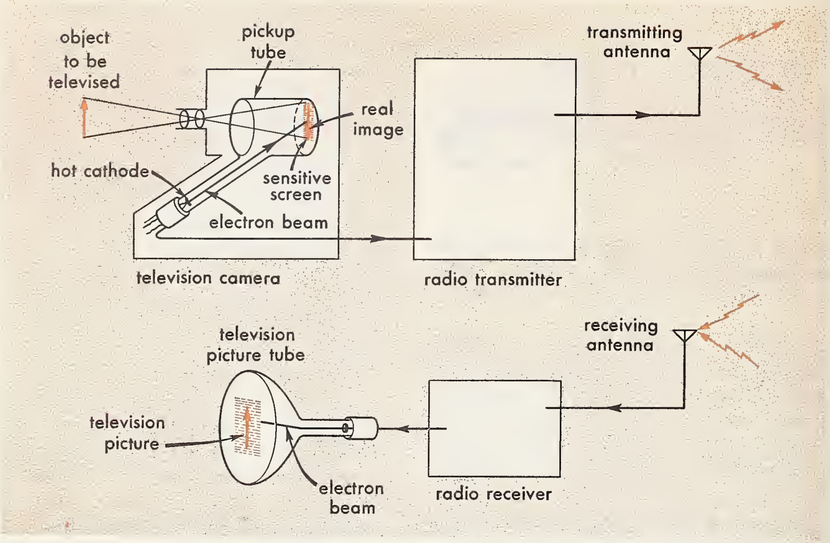


Fig. 32-12. Television transmission and reception.

by the surface atoms of the cesium causes electrons to be ejected from the atoms into the surrounding space. The stronger the light, the greater is the number of electrons emitted.

One of the common uses of a photoelectric cell is as a light-operated electric switch. Such switches can be used to lower the beam of a car's headlights when approaching another car, or they can be used to turn on street lights when it gets dark. Moreover the photoelectric cell is an important part of sound movie projectors and television camera tubes. In a sound movie projector a beam of light passes through the sound track at the edge of the film. This sound track varies the amount of light that reaches the photocell as the film moves past the light, in exact duplication of the sound variations that accompanied the pictures on the film when it was made. This changing light beam then falls on the light-sensitive surface of the cathode of the photocell and causes it to vary in the rate at which it gives off electrons, thus producing a varying voice current which is amplified and fed to loud-speakers.

The Iconoscope. It is impossible to transmit a whole picture by television all in one instant. The picture must be taken apart, so to speak, transmitted bit by bit, and then put back together again at the receiving end. A form of photoelectric cell called an ***iconoscope*** is used to dissect the picture to be transmitted. A diagram of an iconoscope is shown in Fig. 32-12. The electron beam is obtained from a hot cathode in a vacuum tube, the electrons being accelerated and focused into a narrow beam by a high voltage. This beam can be made to sweep back and forth and up and down by applying varying magnetic fields to the beam from ***deflection coils*** mounted around the neck close to the tube.

How is it possible to move the electron beam with a magnetic field? The action is the same as that which takes place in an electric motor, where a flow of electrons in a wire placed in a magnetic field results in a force which moves the wire (the left-hand motor rule). The same things happens in the iconoscope except that the electrons, instead of flowing in a wire, flow through a vacuum in a narrow beam but are nonetheless deflected by a magnetic field. The deflection coils that provide the magnetic field are electromagnets supplied with a varying current, so that a varying force acts on the beam to deflect it back and forth and up and down. When the electrons in the beam reach the end of the tube, they strike a ***mosaic*** screen which is made up of hundreds of tiny separate photocell cathodes insulated from each other and mounted on a metal plate electrically connected to an amplifier. The image of the scene to be televised is focused by lenses onto the mosaic screen (Fig. 32-12). Let us now see how this image is taken apart and transmitted by the iconoscope tube.

The varying magnetic fields of the deflection coils cause the electron beam to strike the upper left-hand corner of the image and sweep across it horizontally to the right-hand edge, then the beam drops a little bit lower and sweeps back again. This sweeping process is repeated many times, each time a little lower until the whole image has been swept or scanned somewhat in the manner in which your eye in reading this page covers the whole page by sweeping across, one line at a time. Each of the tiny photo cathodes on the mosaic screen receives, in its turn, electrons from the beam as it scans the screen. Those cells subjected to very strong light, from the brightest parts of the image formed on the screen, lose a large number of electrons due to the action of the light and so these cells pick up more electrons from the beam than those cells which are subject to only a small amount of light, in the darker parts of the image. Thus as the beam scans the picture a varying current flows into the cells. This varying ***video current,*** as it is called, is amplified many times and then used to modulate a radio transmitter.

The Cathode-ray Tube. How is a television picture received and put back together to form a picture on a screen? The heart of a television receiver is a ***cathode-ray tube,*** commonly called a picture tube. First the radio signals from the television transmitter are picked up by an ordinary radio receiver. The amplified video signal from the receiver is fed to the cathode-ray tube instead of the loud-speaker used in an ordinary radio receiver.

A cathode-ray tube is similar to an iconoscope except that the electron beam falls on a fluorescent coating on the inside of the large end of the tube (Fig. 32-12). The final picture appears on this screen. The fluorescent material absorbs the energy of the high-speed electrons that strike it and some of this energy, instead of being changed into heat, is given off by the atoms of the ***phosphor,*** as the fluorescent material is called, in the form of visible light. The more intense the electron beam, the greater is the intensity of the light given off by the screen at the point where the beam strikes it. Deflection coils cause the beam to scan the screen exactly in step with the electron beam that is scanning the image in the iconoscope of the distant television transmitter. A grid, like the grid in an ordinary triode tube is located near the cathode of the picture tube. This grid is fed with the video signal from the television receiver and varies the ***intensity*** of the electron beam as it scans the screen and thus the picture that was taken apart, bit by bit, by the iconoscope is now put back together on the face of the cathode-ray tube.

Special signals must be transmitted over the air along with the video signal to properly synchronize the sweeping of the cathode-ray tube beam with that of the iconoscope at the transmitter. The sound signals accompanying the picture are carried by radio waves transmitted on a slightly different frequency from the video signals.

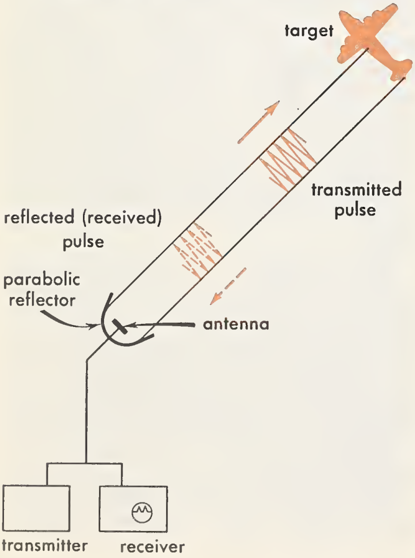
The scanning process takes place at an incredible speed. The beam sweeps back and forth across the screen over 10,000 times per second, taking about $\frac{1}{20}$ of a second to scan the whole image. Because our eyes react so slowly, we are not aware of the actual motion of the beam but see only the completed moving picture that is the end result of the scanning process.

Color television is produced in the same way as black-and-white television, except that the three colors, red, blue and green are transmitted and received separately and then recombined on the face of the picture tube.

Radar. Radar (an abbreviation for ***ra***dio ***d***irection finding ***a***nd ***r***anging) is a device that employs radio waves of very high frequency (100 million cycles per second to 10 billion cycles per second) to locate distant objects through dense fog and clouds, either in daylight or in darkness.

Every radar set contains both a transmitter and receiver. The transmitter sends out a series of short pulses of radio waves focused into a narrow beam by a parabolic reflector placed behind the antenna. The antenna is at the focus of the parabola, and the radio waves are beamed in exactly the same way as electromagnetic light waves are beamed by the reflector in a searchlight or flashlight. The pulses travel out at the speed of light to the target, and are reflected back by it to the antenna where they are picked up by the receiver (Fig. 50-13). The radar antenna rotates, scanning the horizon or the sky. The target is located when a reflected signal is picked up on the receiver. The direction in which the beam is then pointing shows the direction of the target. The distance of the target is found by electronic circuits that automatically measure the time it takes the pulses to reach the target and return.

Fig. 32-13. A radar set receives reflected radio signals from the target it has located.



The apparatus needed for radar is complex and expensive. During World War II radar was used to detect enemy airplanes more than a hundred miles away, even through fog and cloud. Modern military radars are able automatically to aim antiaircraft guns on the target and fire them, or automatically steer a guided missile to its target. Radar is also useful in peacetime in navigating ships and planes through fog and clouds. Shore-

Fig. 32-14. A large radar antenna on the top of a ship's mast.

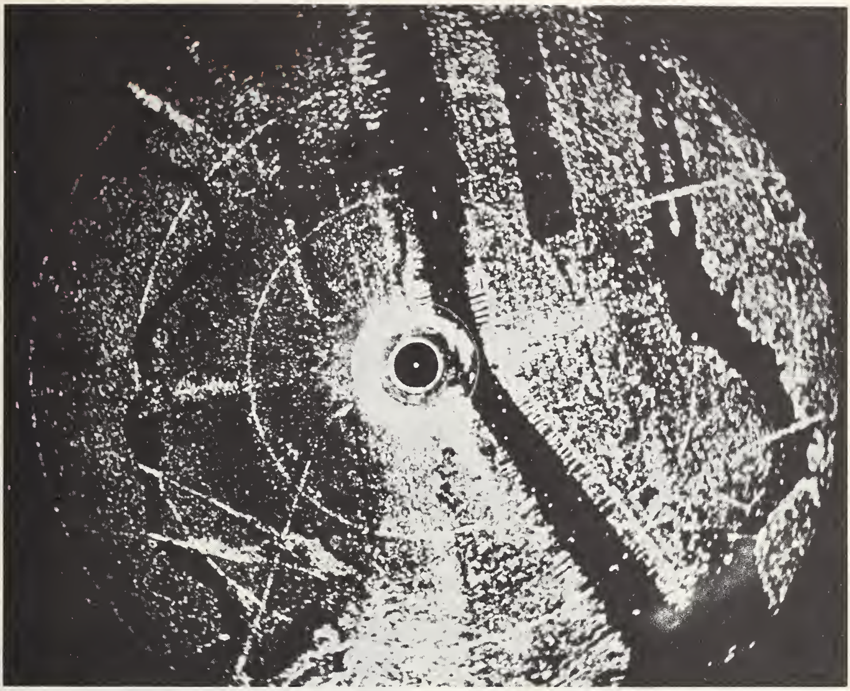


Fig. 32-15. Radarscope photograph of part of New York City and northeastern New Jersey taken from a U.S. Air Force plane. When the photograph was taken, the plane was almost directly over Union City, New Jersey (the spot in the center of the circle). Manhattan Island with Central Park in the northern section and the docks along the river, the Hudson River, the East River, and the Queensboro and Williamsburg bridges joining Manhattan and Long Island are all clearly visible in the right half of the photograph. In the left half, the crowded industrial centers of northeastern New Jersey and the winding course of the Hackensack River can easily be identified. Distance is indicated by the concentric circles used for navigation and bombing.

lines, buoys, icebergs and other ships are easily identified and their positions determined by trained radar operators. Since the very-high-frequency radio waves of radar are reflected by raindrops and snow, radar also serves as a means of locating storms and following their progress.

The type of radar used on the bridge of a ship for navigation is called a PPI (plan, position, indicator) in which a complete, scaled map of the area surrounding the ship appears on the screen of a cathode-ray tube. Such radars are also used to keep track of planes arriving at and leaving airports. The remarkable radar picture shown in Fig. 32-15 is what appears on the cathode-ray tube of the radar set in a plane flying over New York City.

Radio-telescopes. A radar antenna with its curved reflector (Fig. 32-14) suggests the curved mirror used in a reflecting astronomical telescope for gathering light from the stars at night. And in actual fact, a radar antenna and its parabolic reflector, when connected to a very sensitive radio receiver, can be used as a ***radio-telescope*** to pick up radio waves coming from the sun, planets, stars and distant galaxies. If one listens to the radio waves

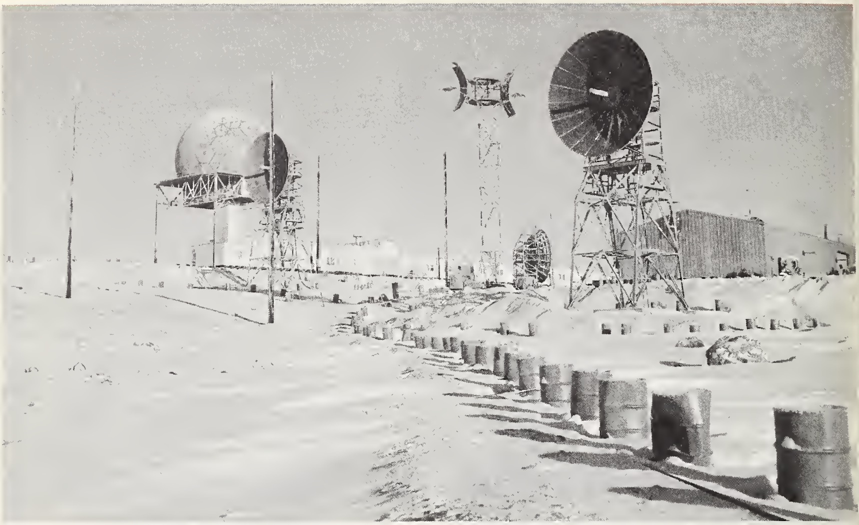


Fig. 32-16. This "Dew (Distant Early Warning) Line" station in Canada is part of the complex radar system intended to give warning of an air attack.

from outer space they sound like noise—hisses and sometimes crackles. Radio astronomers have discovered intense sources of radio waves in the sky where optical telescopes show no stars at all. Static-like radio waves from Jupiter are believed to come from vast storms in Jupiter's atmosphere. Some radio radiations that are picked up are believed to result from the "collision" of two mighty galaxies in distant space, millions of light years from our planet.

Although radio-telescopes are quite new, they are rapidly providing us with a great deal of information about the universe that cannot be obtained with the conventional optical telescope.

Other Uses of Radio and Electronics. In the last two chapters we have only scratched the surface of the capabilities and uses of radio and electronic devices. Other uses include large "electronic brain" computing machines containing thousands of vacuum tubes, automatic navigation of ships and aircraft, proximity fuses for antiaircraft and artillery shells,

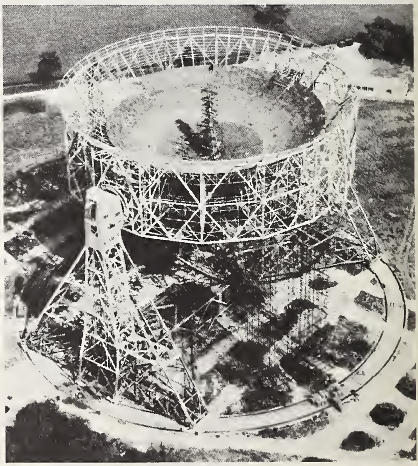


Fig. 32-17. A radio telescope in England that focuses radio waves from galaxies a billion light-years away. These radio waves from outer space give us information about the universe that we are unable to get from light waves and optical telescopes. The building and trees in the upper right-hand corner give some idea of the mammoth size of this telescope.

air conditioning systems, automatic remote weather stations, radiosondes and a host of other applications.

It is remarkable that since Maxwell first predicted the existence of radio waves a century ago, we have come to depend upon them for safety at sea and in the air, for entertainment in the home, and for innumerable conveniences that are provided by various electronic devices.

Things to Remember

Radio waves are electromagnetic waves produced by high-frequency alternating currents flowing in an antenna.

Vacuum tubes are used to generate and amplify radio-frequency current.

In ***amplitude modulation,*** (AM), the radio waves are varied in intensity according to the sound variation to be broadcast.

Modern radio receivers include ***a tuned circuit,*** a ***detector, amplifiers*** and a ***reproducer.***

The ***tuned circuit*** selects the station desired.

The ***detector*** rectifies the radio-frequency alternating currents.

Radio waves can travel long distances when they are reflected around the curve of the earth by the ***ionosphere.***

When light falls on some metals, electrons are released from the metal. This is called the ***photoelectric effect.***

In a ***television system,*** a picture is "taken apart" by scanning with an electron beam and then "put together" on the face of the receiver picture tube.

Radar means, radio direction finding and ranging.

A ***radar*** is a powerful "searchlight" that penetrates fog and haze by using radio waves instead of light waves.

A ***radiotelescope*** focuses radio waves from planets, stars and galaxies.

Questions

GROUP A

1. How were radio waves first generated and detected?
2. How are radio waves generated in modern transmitters?
3. What is the purpose of a radio transmitting antenna?
4. What are the essential parts of a radio receiver?
5. What is the reason for using amplifiers in a radio receiver?
6. How are radio waves able to travel around the curved surface of the earth?
7. What is a photocell and how does it work?
8. How many possible uses of radar can you name?
9. What is a radio-telescope?

GROUP B

10. How is amplitude modulation accomplished?
11. What is the difference between an audio amplifier and a radio-frequency amplifier?

12. Draw a block diagram of an amplitude-modulated transmitter.
13. Why are stations from all over the world heard by a short-wave receiver but not by an FM or TV receiver?
14. What is the purpose of the tuned circuit in a receiver?
15. Why must a rectifier be used in even a simple radio receiver?
16. What is an iconoscope and how does it work?
17. How is a picture "taken apart" and "put together" again in a television system?
18. How does a radar set determine (*a*) the direction of a target and (*b*) the distance of a target?
19. How does the use of a radio-telescope add to our knowledge of the universe?

Things to Do

To build a crystal radio receiver · To build a crystal radio receiver, obtain from a radio service shop a coil, variable condenser, 0.01 microfarad fixed condenser, a 1N21 or similar crystal, and a pair of headphones. The coil and variable condenser from an old worn-out radio will serve the purpose. Connect the apparatus as shown in Fig. 32-18. The antenna should be as long as possible and can be made of any kind of wire, insulated or uninsulated. Since most variable condensers used in radios today are made of two or three separate condensers constructed as a single unit, be sure to use only one section of the condenser. The service man in the shop will probably give you advice and answer your questions about connecting the apparatus. If there is a "ham" (owner of an amateur radio station) in your town, he will be glad to help you.

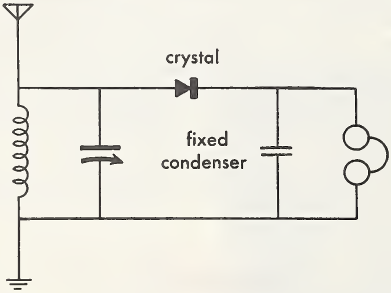


Fig. 32-18. A simple crystal receiver circuit.

If you are interested in building an amplifier for your set or in making a better receiver, write to the American Radio Relay League, West Hartford, Connecticut, for information.

In the catalogues of radio equipment mail-order houses, you will find many "build it yourself" kits listed for sale. These include simple radios, amplifiers, photoelectric cell devices and test instruments.

Atoms and Stars

LOOKING AHEAD

Radioactivity was discovered by accident at the turn of the present century. This accidental discovery has led to profound changes in our way of life; changes in technology, in industry, in medical techniques, and even in the manner of waging war. For better or for worse, we have split the atom.

For a quarter of a century scientists were intrigued by the strange behavior of a radioactive substance such as uranium. It disintegrates and, in doing so, shoots off alpha and beta particles, and gamma rays. This nuclear reaction cannot be controlled; scientists can neither accelerate nor retard its rate of disintegration. Then came a major break. It was found that a single neutron can split the nuclei of a large number of uranium atoms, thereby releasing a vast amount of energy. This is a fission reaction which is controllable. Indeed, the harnessing of atomic power in a reactor may well be the outstanding scientific achievement of the 20th century.

Atomic power promises to compensate for the dwindling reserves of coal and oil. Estimates of available reserves differ, but even optimistic predictions maintain that profitable mining of fossil fuels cannot continue for more than a century. First oil will run out, and then coal. However, the known reserves of fissionable material are many times that of coal and oil together. Herein lies the energy to turn the wheels of industry for centuries to come, and to raise the living standards of backward nations.

Although, on the earth, nuclear fuels are new, they have "burned" in the universe since the beginning of time. But the "fires" of stars are due to fusion, not fission. Nuclei of hydrogen atoms fuse into helium. Vast amounts of heat energy are released in this nuclear reaction, enough to maintain the interiors of stars at unbelievably high temperatures. Indeed, the astrophysicists have computed the internal temperature of the sun to be about 20 million degrees centigrade. And Sirius, the brightest star in the sky, is even hotter than that!

←——— *Stars and luminous dust. Old stars burn out and are continuously replaced by new ones made from dust. (The four spikes surrounding some stars are caused by diffraction of starlight around photographic plate supports.)* (Courtesy of Mount Wilson and Palomar Observatories)

33

POWER FROM ATOMS

The possibility that mass could be converted into energy was first suggested by Einstein more than 50 years ago, in 1905 to be exact. We are all familiar with his well-known equation $E = mc^2$ which, stated in words, is

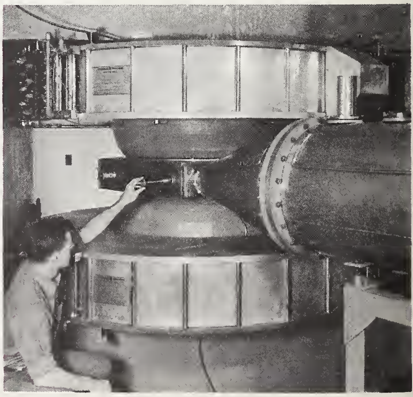
$$\text{energy} = \text{mass} \times \text{velocity of light squared}$$

What does the equation mean? It may be a little difficult to understand the units in which the above terms are expressed, but it actually tells us that a fantastically large amount of energy would be produced by the conversion of a small amount of mass. Or, to be more specific, the Einstein equation says that if one pound of mass could be completely converted to energy, the amount released would be about the same as the energy liberated by burning one and a half million tons of coal. This is hard to believe.

How Can Mass Be Converted into Energy? The large-scale release of energy from mass became the scientist's dream. It gave promise of a fuller life for rich and poor nations alike. If atomic energy could be used as a substitute for oil and coal, even backward countries could be industrialized and their standard of living raised.

But how, as a practical matter, could mass be converted into energy? This was the vital question. One possibility was to bombard the nuclei of atoms with various particles such as protons Ⓟ, alpha particles $\binom{2P}{2N}$, and deuterons $\binom{P}{N}$, which are the nuclei of heavy hydrogen (or deuterium) atoms. Notice, however, that all the above particles are positively charged. Moreover, all atomic nuclei are also positively charged. Why? Hence

Fig. 33-1. A cyclotron. This picture shows the two poles of an electromagnet, above and below the vacuum tank where the operator is making an adjustment. A rapidly changing electric field within the tank causes charged particles to whirl at high speed. The high speed particles escape from the tank through the tube at the right. In doing so, they strike a target and smash the nuclei of some of its atoms.



there is an electrical force of repulsion between the particles and atomic nuclei. Therefore, to hit a nucleus (that is, to overcome the force of repulsion) the particles had to be hurled at high speeds, speeds approaching the velocity of light. Various atom smashers were devised for this purpose, the most familiar being the cyclotron invented by an American physicist, Ernest Lawrence.

Artificial Radioactive Isotopes. Atom smashers did not turn out to be efficient machines for producing atomic energy but they enabled scientists to make artificial radioactive isotopes. Lawrence, for example, made radioactive sodium by hurling deuterons against a target of common salt. If a deuteron managed to hit the nucleus of a sodium atom $\begin{pmatrix}11\text{P}\\12\text{N}\end{pmatrix}$ a violent upheaval took place in which the deuteron split into a proton and a neutron. The neutron was captured by the nucleus, but the proton was repelled and ejected. The nuclear change can be shown as follows:

$$\begin{pmatrix}11\text{P}\\12\text{N}\end{pmatrix} + \begin{pmatrix}\text{P}\\\text{N}\end{pmatrix} \rightarrow \begin{pmatrix}11\text{P}\\13\text{N}\end{pmatrix} + \text{P}$$

What is $\begin{pmatrix}11\text{P}\\13\text{N}\end{pmatrix}$? This nucleus has the same number of protons as the starting substance, sodium-23. Hence it is an isotope of sodium, sodium-24. However, the nucleus of sodium-24 is unstable and, to achieve stability, one of its neutrons changes to a proton and a beta particle (an electron), which is ejected. In other words, sodium-24 is radioactive, and it disintegrates as follows:

$$\begin{pmatrix}11\text{P}\\13\text{N}\end{pmatrix} \rightarrow \begin{pmatrix}12\text{P}\\12\text{N}\end{pmatrix} + \text{electron}$$

The nucleus $\begin{pmatrix}12\text{P}\\12\text{N}\end{pmatrix}$ shows that an entirely new element has been formed in the radioactive disintegration. Why? This new element is magnesium, which is stable. These changes can be summarized as follows:

$$\text{sodium-23} + \text{deuteron} \rightarrow \text{sodium-24} + \text{proton}$$

$$\text{sodium-24} \rightarrow \text{magnesium-24} + \text{electron}$$

A great many radioactive isotopes were made from light elements in a cyclotron. Indeed, an entirely new field of radioactive synthetic elements was opened up which offered great promise in industry, agriculture, and medicine. But what of the release of energy from atoms? When a bulls'-eye was scored on a nucleus, energy was certainly released. But the chance of a collision was so remote that far more energy was expended in hurling the atomic projectiles than was released when nuclei were hit. Was it possible to use projectiles that required less expenditure of energy, projectiles that need not be hurled by a cyclotron? The answer to this question suggested the use of neutrons.

Neutrons as Bullets. As already stated, protons and alpha particles are repelled by a nucleus; neutrons, on the other hand, having no charge, are not repelled. Instead, they move freely through the electrons surrounding the nucleus and into the nucleus itself. It was the Italian-born American physicist, Enrico Fermi, who in 1934 first used neutrons as projectiles. Indeed, it was Fermi's work on neutron bombardment that turned out to be the key that opened the door to atomic energy.

What did Fermi use as a neutron source? Neutrons cannot be easily dislodged from a nucleus and Fermi obtained them by an indirect method. He discovered that if alpha particles from radium strike the nuclei of beryllium atoms neutrons are ejected. The reaction is

$$\begin{pmatrix}4\text{P}\\5\text{N}\end{pmatrix} + \begin{pmatrix}2\text{P}\\2\text{N}\end{pmatrix} \rightarrow \begin{pmatrix}6\text{P}\\6\text{N}\end{pmatrix} + \text{N}$$

$$\text{beryllium} + \underset{\text{particle}}{\text{alpha}} \rightarrow \text{carbon} + \text{neutron}$$

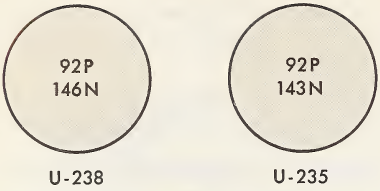


Fig. 33-2. The two isotopes of uranium are U-238 (plentiful) and U-235 (rare).

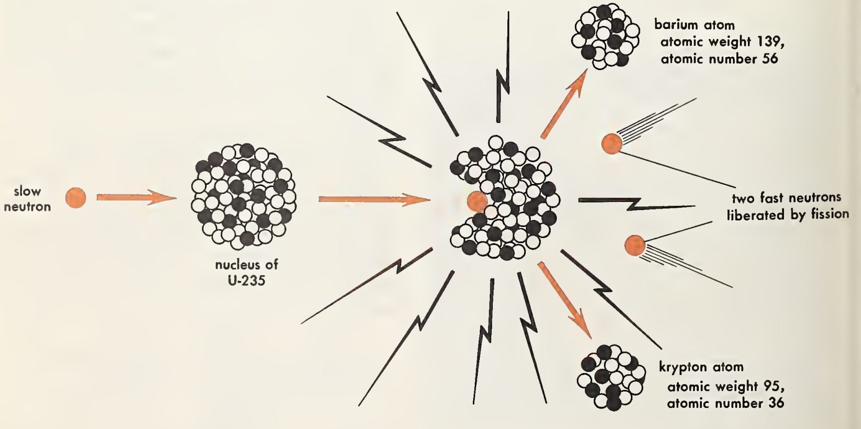
That is, neutrons are produced artificially in this reaction, and it probably played a part in the explosion of the first atom bomb.

Bombardment of Uranium by Neutrons. What would happen if neutrons were used to bombard heavy elements like uranium, elements whose nuclei are unstable and radioactive? Would the nuclei become still more unstable if they were penetrated by a neutron? The bombardment of uranium by neutrons gave some unexpected results—the particular effect is different for each of the isotopes of uranium. The plentiful isotope (uranium-238) absorbs slow neutrons to form a synthetic element, plutonium, which is heavier than uranium. The rare isotope (uranium-235) absorbs a slow neutron and literally blows itself apart. The nucleus of U-235 splits into two main fragments, a process called ***fission,*** releasing a vast amount of energy so that the fragments fly apart with great violence. This discovery was made in the early part of 1939 by a German physicist named Otto Hahn, a few months before the outbreak of the Second World War. The advent of the war raised the same question among scientists, whether friend or foe, Could the fission reaction be used to make an atomic bomb?

Fission. ***Fission is the process in which a nucleus breaks up into two fragments of approximately equal size, releasing neutrons, gamma rays, and a great deal of heat energy.*** This release of energy occurs because the fragments weigh a little bit less than the original mass. Suppose, for example, that one pound of U-235 is fissioned. The loss of weight would amount to only 1/1000 pound. But according to the Einstein equation, this minute loss of mass results in the release of a vast amount of energy.

We must not think, however, that the loss of mass is caused by the annihilation

Fig. 33-3. U-235 can be fissioned by slow neutrons. The fission products must account for the total mass except for the small amount of mass converted to energy (1+235=139+95+2). Also the charge is conserved (92 protons=56 protons+36 protons).



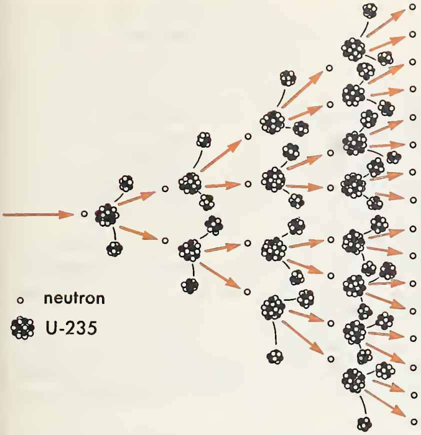


Fig. 33-4. A chain reaction. If each fission of U-235 released 2 neutrons, the chain reaction will be almost instantaneous.

of protons or neutrons—there are exactly the same number after fission as before, but the nuclei of the fragments are packed differently. Indeed, the fragments contain so many neutrons that they are unstable. To achieve stability several things can happen to a nucleus; some neutrons are emitted; some neutrons are converted to protons, and electrons are emitted. The fragments are therefore strongly radioactive, emitting beta particles or neutrons and possibly gamma rays.

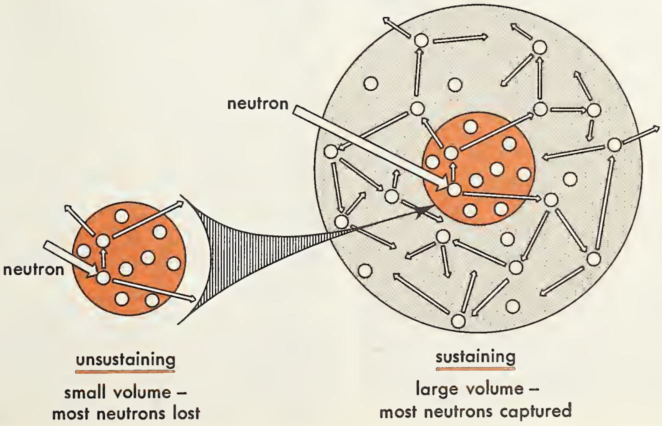
A Chain Reaction. Since neutrons are released when a nucleus of U-235 is fissioned, scientists wondered if a ***chain reaction*** were possible. Or, in other words, would these released neutrons penetrate the nuclei of other atoms so that the entire mass would explode in very short time? We now know that such a chain reaction does, in fact, take place.

Suppose we assume that, on the average, two neutrons are released per fission. The 2 neutrons released in the first fission then attack two other nuclei and release 4 neutrons; the 4 neutrons would attack 4 other nuclei and release 8; the 8 neutrons would release 16, and 16 would release 32, and so on until no fissionable material is left. Actually a chain reaction takes place in an exceedingly short period of time, less than a millionth of a second. It is apparent that a reaction of this kind could lead to a catastrophic explosion.

An Atomic Bomb. Before a bomb could be made it was necessary to separate the two isotopes, U-235 and U-238, an exceedingly difficult operation.

The next complication is that of ***criti-***

Fig. 33-5. If the uranium is less than critical size (as on the left), too many neutrons escape into the air, and a chain reaction cannot get started.


neutron
neutron
unsustaining
small volume – most neutrons lost
sustaining
large volume – most neutrons captured

cal mass. If the mass of U-235 is very small a large number of neutrons released by fission escape through the surface, and a chain reaction is prevented. (See Fig. 33-5.) If a larger mass is taken, relatively fewer neutrons escape into the air; if the mass is the right size (the critical mass), enough neutrons are kept within the mass to set up a chain reaction and the U-235 instantly explodes.

Details of the mechanism used to "fire" a bomb have never been published, but it is clear that there are at least two ways to set off the explosion. In the safety position, the uranium mass is separated into two or three parts, each less than critical size. If the total mass is greater than critical mass, it will explode instantly if the parts are brought rapidly together. One device for driving the pieces together might be some form of gun. Another might depend on an explosion. Suppose, for example, that a core of subcritical pieces of U-235 (or plutonium) is surrounded by some chemical explosive which is detonated. The high gas pressure caused by the explosion would be sufficient to compress the pieces and to increase the core to critical mass.

When fission occurs, the atomic fragments are forced apart at tremendously high speeds. As they collide with each other their energy of motion is changed to heat energy. The temperature in the bomb at the instant of fission is believed to be at least 10 million degrees centigrade, about the same as the temperature of the interior of the sun. At this temperature all reaction products are gases; moreover the surrounding air is greatly expanded. This sudden expansion creates a shock wave which causes severe destruction. In addition, the highly penetrating gamma rays, emitted by the radioactive products, are highly dangerous. Indeed, the radioactivity emitted by a single bomb is equivalent to that emitted by almost 100 tons of radium. Finally, there is contamination over a wide area from radioactive ***fall-out,*** a subject discussed on page 345.

The first atomic bomb, the outcome of almost five years of intensive research, was an experimental one exploded in a New Mexico desert in July, 1945. The energy released in this explosion was equivalent to the energy liberated by exploding 20,000 tons of TNT. The following month bombs were dropped on Hiroshima and Nagasaki. The bomb dropped on Hiroshima was made of uranium-235, that dropped on Nagasaki, three days later, was made of plutonium. The effects were catastrophic. In Hiroshima, for example, 70,000 people were killed, 300,000 wounded, and most buildings within a radius of three miles were destroyed.

And so ended the Second World War; the atomic age was born in violence. But the same forces that wreaked destruction in war can be harnessed for the benefit of mankind. The machine which diverts the energy of fission to useful purposes is called a ***reactor.***

A Uranium Reactor. The rate of fission in a reactor must be controlled so that it is much slower than in a bomb. Fermi suggested that it might be possible to use ***natural*** uranium as a fuel, thus avoiding the laborious separation of the isotopes. The neutrons must be slowed down so that they could be more easily captured by U-235, which is present in such small proportion. A substance which slows down neutrons is called a ***moderator.*** Fermi used graphite as a moderator.

Fermi built the first atomic reactor in 1942 under the stands of the athletic field of the University of Chicago. It consisted of cubic blocks of graphite in which were placed lumps of natural uranium and uranium oxide. In all, 50 tons of uranium and 500 tons of graphite were enclosed in a six-foot wall of concrete to prevent the escape of dangerous radiations and neutrons.

As U-235 is fissioned, fast neutrons are released. Some of these are slowed down by graphite so that they can be captured by other nuclei of U-235 and a chain reaction induced. Boron steel rods control

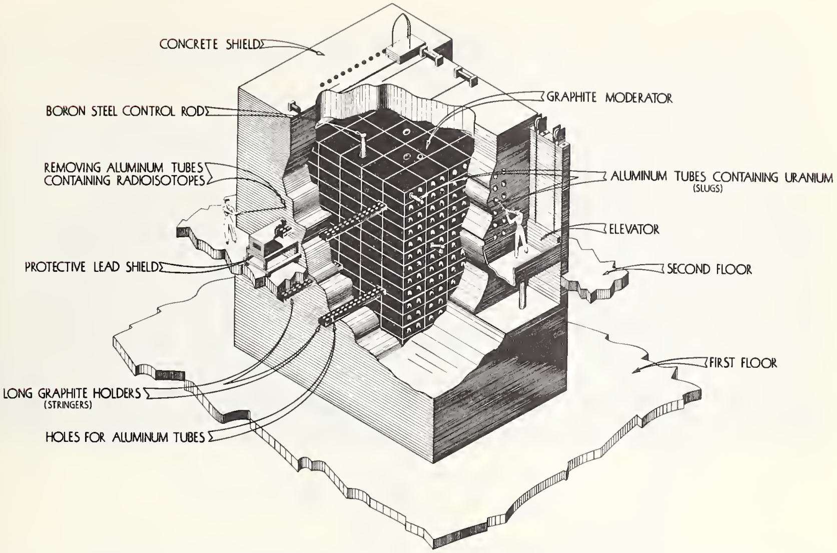


Fig. 33-6. The first reactor used graphite as a moderator and boron steel rods to control the rate of fission. Notice the heavy wall of concrete surrounding the reactor.

the rate of fission and prevent the reaction from getting out of hand. Actually boron is a good absorber of neutrons and, if the rods are pushed into the reactor, the fission process is slowed down and may be stopped altogether. Some of the fast neutrons, on the other hand, are absorbed by U-238 to form the element plutonium.

Here then was a new kind of heat engine that offered great promise. Uranium-235 was fissioned with the release of heat energy. And, at the same time, uranium-238 was changed to plutonium, which provided additional fuel for the reactor. Plutonium, however, is exceedingly toxic, and great care must be exercised in handling it. It emits alpha rays and is 200,000 times more active than uranium. Moreover, like radium, it is very dangerous if it gets within the body, a topic considered on page 343.

Power from Reactors. A reactor is started by withdrawing the boron safety rods. As the chain reaction develops, the rods are gradually inserted to maintain activity at the desired level. The interior of a reactor becomes intensely hot. In the early reactors (similar to that shown in Fig. 33-6), scientists were only interested in making plutonium or radioactive isotopes, and the heat was wasted. Actually the heat was carried away by enormous fans which blew about 100,000 cubic feet of air per minute through the channels of the reactor.

Modern reactors are, however, important heat engines, and engineers are learning how to conserve the heat and use it to generate electricity. Although one pound of U-235 consumed in a reactor produces as much heat as 1500 tons of coal burned in a furnace, there are many problems to be solved before uranium can profitably replace coal in the production of electricity. First of all, reactors are very expensive; they cost from 25 to 50 million dollars, depending on their size. Then again structural metals, steel for example, corrode and lose their strength when bombarded by neutrons and radioactive particles at high tempera-

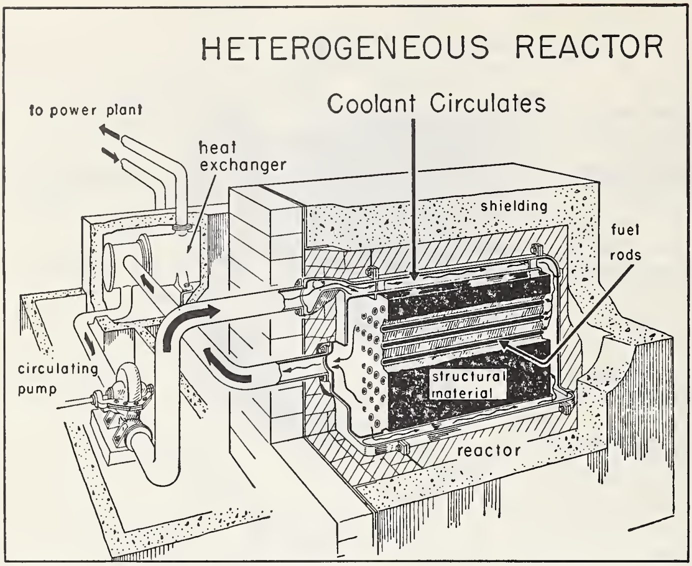


Fig. 33-7. A modern reactor. Notice particularly the channels, between the uranium rods and the moderator, through which the coolant flows. The heat of the coolant is transferred to water at a heat exchanger. Steam from the water is used to generate electricity.

tures. However, this problem is being solved by using "new" metals in the construction of reactors; zirconium, for instance, absorbs few neutrons and is highly resistant to corrosion. Finally, there is the disposal of radioactive wastes.

A reactor, like a bomb, has a critical size. If natural uranium is used, a cubic reactor may have sides measuring 40 feet or so. If, however, the fuel is enriched, so that the proportion of U-235 to U-238 is increased, a much smaller reactor will provide enough neutrons to keep the chain reactor going.

A possible design for a reactor is shown in Fig. 33-7. Rods of enriched uranium are embedded in a moderator. Between the rods and the moderator is a channel through which a liquid coolant, such as molten sodium, can circulate and carry off the heat. The hot fluid coolant leaves the reactor, enters a heat exchanger, and then returns to the reactor. The coolant picks up neutrons in the reactor and becomes highly radioactive. For this reason, coolant and water are on different circulating systems and, by means of a heat exchanger, heat is transferred indirectly from coolant to water.

A recent experimental boiling-water reactor even eliminates the heat-exchanger. In this new type, live steam is generated in the core of the reactor and the steam is piped directly to a turbine.

Nuclear Power-Driven Ships. Atomic reactors are now being used to power ships at sea, particularly submarines. The advantage of reactor-driven submarines is obvious; their "engines" require no oxygen and they can remain submerged indefinitely. Moreover, uranium is a potent fuel; it enables a submarine to operate continuously for many months without the need of refuelling.

The first atom-powered submarine, the *Nautilus*, was launched in 1954. Heat from its reactor is transferred to water by a heat exchanger. The water is changed

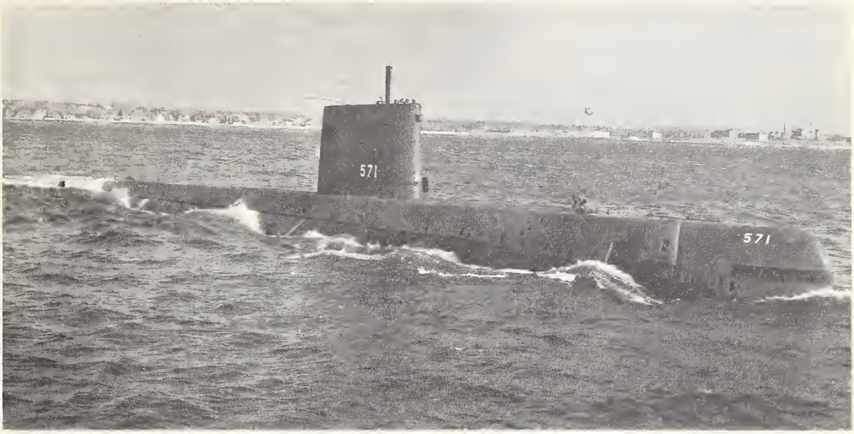
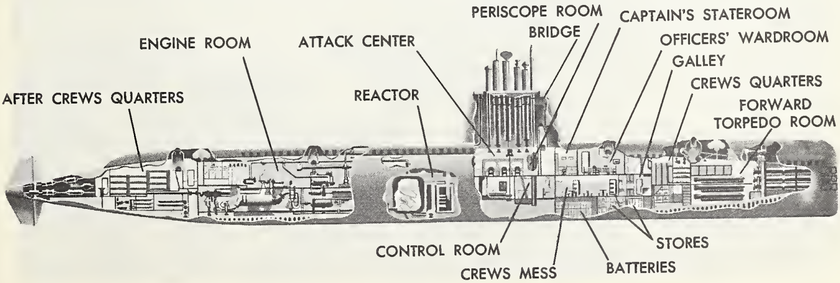


Fig. 33-8. *U.S.S. Nautilus,* the first ship driven by power from atoms. In its first cruise the *Nautilus* travelled more than 60,000 nautical miles without refueling.

to steam, and the steam drives turbines which, in turn, rotate propellers. The crew are protected against harmful radiations by a lead shield which surrounds the reactor. The *Nautilus* travelled more than 60,000 nautical miles without refuelling. A similar cruise by a diesel-powered submarine would have required about 720,000 gallons of oil. The *Nautilus,* however, "consumed" a charge of U-235 no larger than an ordinary light bulb.

Radiation Effects. How dangerous are the radiations (gamma rays, neutrons, alpha and beta particles) thrown off by fission products in a reactor or a bomb explosion? Gamma rays, which are even more powerful than X rays, penetrate deeply into human tissue; and so do neutrons. On the other hand, neither alpha nor beta particles can penetrate deeply. However, if the atoms that emit these particles get inside the body, for example, during breathing, the radiations can be very dangerous. All radiations tear apart the molecules that make up human tissue; protein molecules, for example, are readily damaged by radiations. As a result, cells may be killed by radiations and various organs in the body may be changed biologically. It is apparent that adequate safety precautions are essential. This is doubly necessary because radiations do not affect the senses; that is to

Fig. 33-9. The reactor on the *Nautilus* supplies power to its engines.



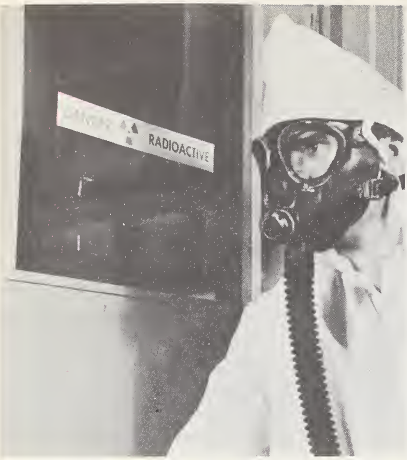


Fig. 33-10. Employees are protected against radiation hazards. This man is about to enter a "hot" room. He wears a laboratory coat, rubber gloves, shoe covers, a head cover, and a breathing mask connected to an outside air supply.

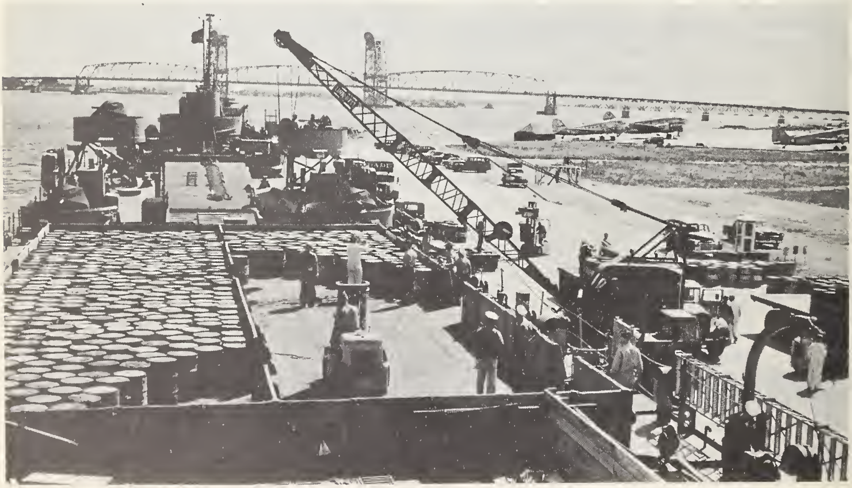
say, the effects of radiation bombardment cannot be felt, and a man could receive a fatal exposure without even being aware of it.

Reactor Hazards. Reactors generate radioactivity in vast quantities; the radioactivity of an average size reactor is equivalent to that of 300 tons of radium. Indeed, unless a reactor were surrounded by a six-foot shield of concrete, all people within a quarter-mile of it would receive a fatal dose within a few minutes. Nevertheless, in spite of the risks, the safety record of reactor employees is almost without parallel in industry.

One question that comes to mind is, Can reactors, like bombs, explode? The answer is "No." Even if all controls failed to operate so that the fission rate greatly increased, the heat developed would probably cause the reactor to break apart. Fission would then cease, because the reactor would be less than critical size. Nonetheless, a broken reactor would be a disaster, unless proper precautions were taken, because highly dangerous radioactive particles would be released into the air.

Radioactive Wastes. In a fission reaction there are many kinds of nuclear fragments. Indeed, as many as 35 different radioactive elements may be formed. These for the most part constitute use-

Fig. 33-11. Disposal of radioactive wastes. These solid radioactive wastes are embedded in concrete and sealed in steel drums. The drums are about to be taken out to sea and dumped in water more than a mile deep.



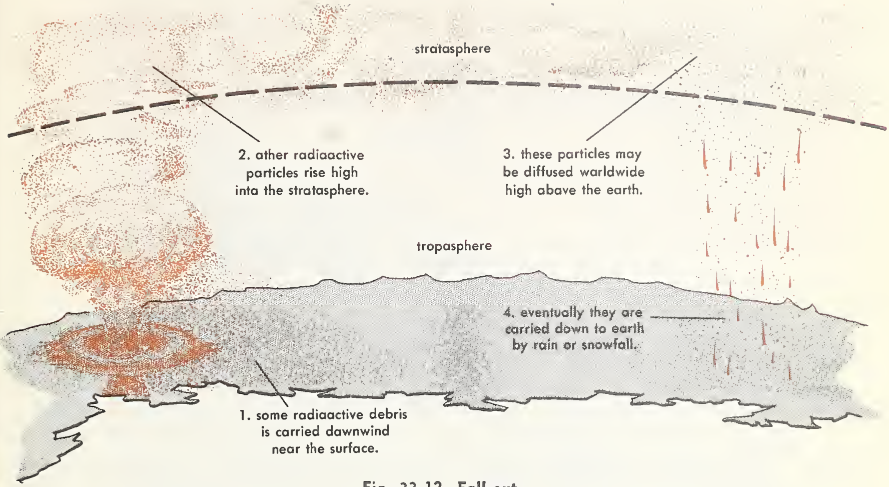

stratosphere
2. other radioactive particles rise high into the stratosphere.
3. these particles may be diffused worldwide high above the earth.
troposphere
4. eventually they are carried down to earth by rain or snowfall.
1. some radioactive debris is carried downwind near the surface.

Fig. 33-12. Fall-out.

less radioactive wastes and must be removed from reactors from time to time, just as we remove ashes from a furnace. Radioactive wastes emit dangerous radiations for months and even for years. How can radioactive wastes be disposed of? So far there is no satisfactory answer to this question. Sometimes they are stored in "burial grounds" far from populated regions; sometimes they are stored in enormous underground tanks until they are no longer dangerously active; sometimes they are sealed in concrete containers and dumped into the sea. It is obvious that as more and more reactors are built the disposal of radioactive wastes will constitute a major problem, one that may be solved by finding beneficial uses for the wastes.

Fall-out. In an atomic bomb explosion a large number of released radioactive elements escape into the atmosphere and are carried away by winds. If the fireball of an explosion touches the ground the radioactive elements adhere to fine particles of rock or dust, forming ***fall-out***. The fall-out is sucked up into the atomic cloud and carried high into the atmosphere. As it slowly settles to earth, it causes radioactive contamination over a wide area. Some fall-out, particularly after a hydrogen bomb explosion, is blown into the stratosphere. Here the particles cannot be washed to earth by rain as in the troposphere, and they may be held up by air currents for years before ultimately settling to earth.

One of the most dangerous constituents of fall-out is radioactive strontium, or strontium-90; that is, the strontium isotope of atomic weight 90. It is one of the fission products of uranium. Strontium-90 has a half-life of 28 years, far longer than the half-lives of most other fission products. Like radium, it is a beta emitter. Also like radium, it settles in the bones if it gets into the body and, if present in sufficient amounts, can induce cancer. As strontium-90 returns to earth it mixes with chemical elements in soil. Vegetables then take up radio strontium along with calcium. If the vegetables are eaten, small amounts of strontium-90 enter the stomach and then pass into the bloodstream. Here lies a risk that could become a threat to all human life if atom bombs play a major role in future warfare.

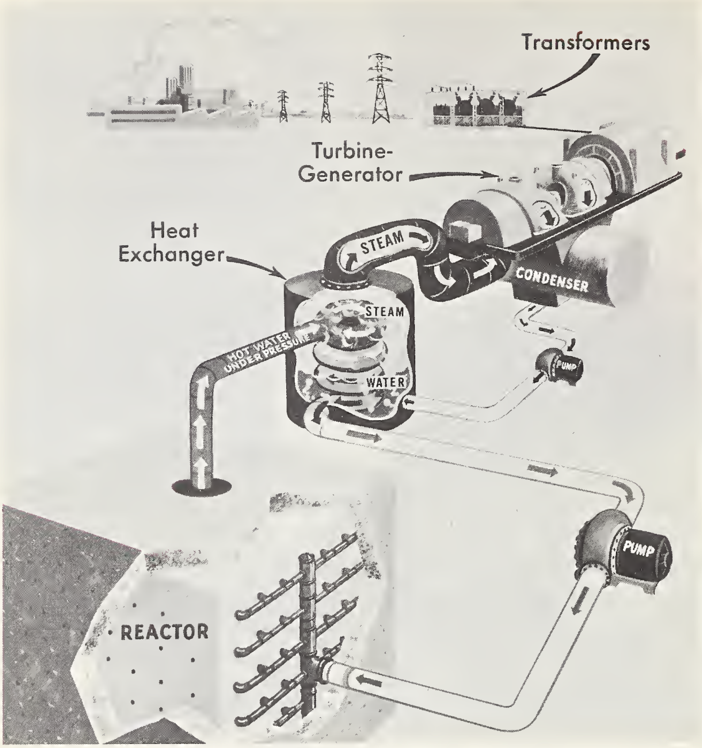


Fig. 33-13. Atomic energy is changed to heat energy in a reactor. This heat can be used to drive a turbine-generator and provide the electricity to operate motors in industrial plants.

Will Atomic Power Displace Coal and Oil? All nations need ample sources of energy; without energy industries cannot be supported, and a nation without industries is committed to a low standard of living. In the past, a nation without coal and oil has been unable to maintain industries on a large scale. But the prospect of power from reactors promises to redress the effects of the uneven distribution of natural resources. Indeed, the discovery of a way to extract energy from atoms has made it possible to bring about profound social changes in backward countries and may turn out to be one of the most significant of all scientific achievements.

In Asia, Africa and South America there are vast regions entirely devoid of coal and oil. Reactors built at strategic places in these areas would stimulate industry and revolutionize their ways of life. In Canada, as we have seen, there is an abundance of coal and oil. It is therefore unlikely that the standard methods of producing power will be abandoned for some years, particularly as power from uranium is more expensive than power from coal or oil. Nevertheless, several experimental reactors have been built in Canada and reactor research continues.

But even if all the useful power in the world were produced by splitting atoms there still would be a great need for coal and oil. Let us remember that coal and

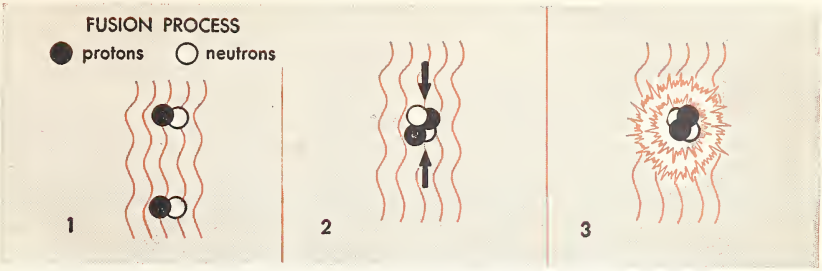


Fig. 33-14. If 2 protons and 2 neutrons (from hydrogen nuclei) make close contact at high speed, they fuse together, forming a helium nucleus and liberating much heat energy.

oil are the raw materials for many chemical industries, some of which have been discussed in earlier chapters.

Fusion. Energy is released when nuclei of heavy atoms like uranium and plutonium split into light ones; energy is also released when the lightest atoms combine to form heavier ones. These latter reactions are called ***fusion*** reactions.

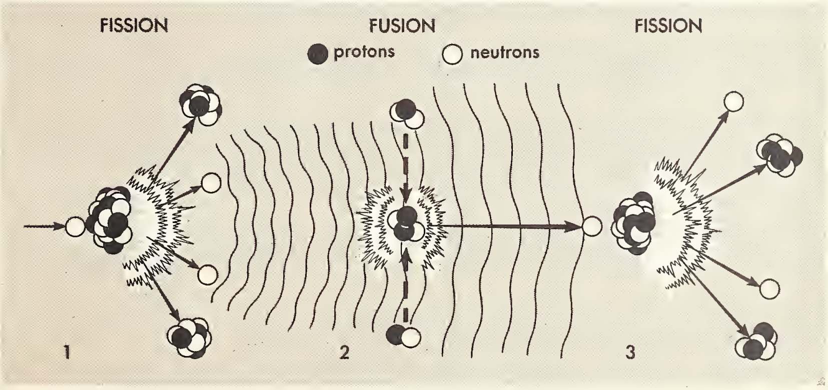
The simplest fusion reaction would be the combination of four hydrogen atoms to form a helium nucleus. Or,

$$4\ \textcircled{P} \rightarrow \textcircled{\begin{matrix}2\text{P}\\2\text{N}\end{matrix}} + 2\ \text{electrons}$$

The weight of the helium nucleus is less than the weight of the reacting protons and electrons, and the mass lost would be converted into energy. Moreover, the conversion of mass to energy is about seven times greater in a fusion reaction than in a fission reaction.

Fission reactions take place readily with slow neutrons, but fusion is a difficult operation to bring about. Why is this? All nuclei are positively charged and, to make them combine they must move at speeds high enough to overcome their strong force of repulsion. For this reason, fusion takes place only at exceedingly high temperatures. Indeed, the only known fusion reactions in nature are in the interior of stars, where the temperatures are at least 20 million degrees centigrade. Thus it is believed that

Fig. 33-15. Hydrogen bomb reactions. The fission reaction on the left produces a temperature high enough to trigger the fusion reaction. Neutrons released in the fusion reaction may be used to trigger another fission reaction.



the sun derives its energy from the fusion of hydrogen to helium atoms.

A Hydrogen Bomb. The fusion of hydrogen atoms would not be a satisfactory source of energy for a hydrogen bomb. The reaction is far too slow; less than one per cent of hydrogen in the sun is converted into helium in a billion years. However, the heavier isotopes of hydrogen (deuterium and tritium) could be used in a "hydrogen" bomb, since they fuse more rapidly than hydrogen and at a lower temperature. This reaction is

deuterium + tritium → helium + neutron

$$\begin{pmatrix}\mathrm{P}\\\mathrm{N}\end{pmatrix} + \begin{pmatrix}\mathrm{P}\\2\mathrm{N}\end{pmatrix} \rightarrow \begin{pmatrix}2\mathrm{P}\\2\mathrm{N}\end{pmatrix} + \mathrm{N}$$

Notice that, since there is no neutron bombardment of a nucleus, the concept of critical size does not arise in fusion reactions; there is no limit to the size of a hydrogen bomb. The combined effects of larger size and the higher conversion of mass to energy together make a hydrogen bomb far more destructive than a fission bomb.

Even less information has been published about the mechanism of a fusion bomb than of a fission bomb. Nevertheless, it is known that the only way to produce the high temperature necessary for fusion is to explode a uranium or plutonium bomb. Thus a hydrogen bomb probably consists of a central exploder (a U-235 bomb) surrounded by lithium-deuteride. Lithium-deuteride is a compound of deuterium and the light isotope of lithium, lithium-6.

Lithium-6, $\begin{pmatrix}3\mathrm{N}\\3\mathrm{P}\end{pmatrix}$, is probably split into tritium and helium by neutrons released in the fission reaction. Or,

$$\begin{pmatrix}3\mathrm{N}\\3\mathrm{P}\end{pmatrix} + \mathrm{N} \rightarrow \begin{pmatrix}\mathrm{P}\\2\mathrm{N}\end{pmatrix} + \begin{pmatrix}2\mathrm{P}\\2\mathrm{N}\end{pmatrix}$$

The tritium then fuses with deuterium with explosive violence. Or,

$$\begin{pmatrix}\mathrm{P}\\2\mathrm{N}\end{pmatrix} + \begin{pmatrix}\mathrm{P}\\\mathrm{N}\end{pmatrix} \rightarrow \begin{pmatrix}2\mathrm{P}\\2\mathrm{N}\end{pmatrix} + \mathrm{N} + \text{energy}$$

In other words, there are probably three different reactions in a hydrogen bomb explosion. The explosive power of such bombs can exceed 20 million tons of TNT, which is 1000 times the explosive power of a standard fission bomb!

Things to Remember

Einstein's equation is ***energy = mass × velocity of light squared.***

Neutrons are good ***projectiles*** since they are not repelled by the charged nuclei of atoms.

Fission is the splitting of a large nucleus into two fragments of approximately equal size.

The minimum mass of uranium (or plutonium) in which a fission chain reaction is possible is called its ***critical mass.***

A ***reactor*** is a machine that converts the energy of fission to useful purposes.

Fission products are radioactive.

The most ***dangerous radiations*** are gamma rays and neutrons.

Fall-out is radioactive material that settles to earth after an atomic bomb explosion.

Strontium-90 is the most dangerous constituent of fall-out.

Fusion is the combination of the nuclei of light atoms.

The ***heat of the sun*** is due to the fusion of hydrogen atoms into helium.

Questions

GROUP A

1. What is meant by the term ***fission?***
2. Explain what happens in a chain reaction.
3. What is the meaning of the term ***critical mass?***
4. What is the purpose of a moderator in a reactor?
5. Why is zirconium used in the construction of reactors?
6. What determines the size of a reactor?
7. What is fall-out?

GROUP B

8. (*a*) State Einstein's equation. (*b*) What does it mean?
9. Explain what happens when sodium-24 changes to magnesium.
10. Explain clearly why U-235 is explosive above its critical mass but not explosive below its critical mass.
11. Explain (*a*) how heat is released in a reactor and (*b*) how this heat can be put to useful purposes.
12. (*a*) What are the effects of radiations on human tissue? (*b*) Why is strontium-90 particularly dangerous?
13. What are radioactive wastes and how are they disposed of?
14. Explain the source of the sun's energy.

Things to Do

Becquerel's experiment · You can repeat Becquerel's experiment if you have a uranium ore. Rest the ore on a key placed on a covered photographic plate in a dark room. Leave the ore on the plate for a week or so and then develop the plate. Notice that the plate is fogged and that there is an impression of the key on it. How can you account for this?

34

THE SUN

THE STUDY of heavenly bodies is known as ***astronomy,*** a word derived from the Greek ***astron,*** a star. Astronomy was the first science to be studied by man. Throughout the ages no heavenly body has fascinated man more than the sun. Its regular appearance on the eastern horizon at daybreak, its course across the sky through the day, and its setting on the western horizon were all observed by early man. We know from records that, thousands of years ago, people worshiped the sun. Probably the oldest records of sun worship in existence are huge stones, weighing many tons, in Brittany, France, and at Stonehenge in England. Their arrangement shows that men of the New Stone Age observed the motions of the sun, and with these observations they measured time.

From the apparent motion of the sun in the sky early man concluded that the earth was fixed and that the sun moved around it. He could not "see" the earth in motion, so there was nothing to convince him that he was on a moving planet. If you are in a train alongside another train at a railroad platform you may find yourself in a similar quandary. Suppose one train starts to move, how can you tell which train is moving? You can, of course, see that one of the trains is in motion but you cannot tell which it is unless you can see a stationary object, such as the platform. If you are not moving with respect to the platform, then clearly it is the other train that moves, not yours. Early man knew, of course, that either the sun or the earth was in motion. But he made a wrong guess, and the false notion that the sun

Fig. 34-1. This obelisk from Alexandria, Egypt, is now in Central Park, New York. The early Egyptians used obelisks as clocks and calendars.

moves around the earth persisted for almost 2000 years.

Egyptian Astronomy. The ancient Egyptians knew many astronomical facts, particularly concerning the sun, moon and planets. Their obelisks, some of their temples, and even some details of the construction of their pyramids, show that their positions are linked to that of the sun at definite times of the year. The obelisks, or shadow clocks, were among the earliest calendars. The shadow of an obelisk moves as the earth travels along its orbit and lengthens or shortens with the seasons. At least three of these ancient towering, four-sided, pointed pillars have been exported from Egypt. One is on the Thames Embankment in London, another in the Place de la Concorde in Paris, and a third, known as Cleopatra's Needle, is in Central Park, New York.

Greek Astronomy. The Greeks had a more poetic concept of the sun than the Egyptians. To them the sun was the fiery golden carriage of the god Apollo racing across the sky. However, it was a Greek philosopher name Aristarchus who proposed the idea that the sun, and not the earth, was the center of the universe. He stated that the earth revolved about the sun in a circular orbit and that the earth was just a ***planet,*** or "traveler," that received its warmth and light from the sun. But few people believed him—he was centuries ahead of his time. The accepted theory was that stated by Aristotle, one of the greatest of the Greek philosophers. Aristotle maintained that the earth was fixed and immovable, and that the sun and all other heavenly bodies moved around it.

Astronomy in the 16th and 17th Centuries. In the 16th century a Polish astronomer named Copernicus revived the ideas of Aristarchus and supported them with observations and reasoning. In his day telescopes had not been invented, and he therefore found another means of observing the heavens. He cut slits in the walls of his house in such positions that he could watch a few planets for successive nights. He noticed, for example, that the brightness of Mars varied considerably and, because of variations in brightness, he could not believe that heavenly bodies remained at fixed distances from the earth. Indeed, he explained the movement of these bodies by assuming that the earth made a complete rotation from west to east every 24 hours in a direction opposite to the apparent motion of the planets. Copernicus therefore rejected the idea that a gigantic universe revolved about the earth. Moreover, he recognized the earth as a part of the sun's system and distinguished between the sun's system and the stars. He produced evidence to show that the sun is the center of the solar system and that the planets revolve about the sun and, at the same time, rotate about their own axes.

Galileo. Early in the 17th century Galileo, the famous Italian scientist, made the first astronomical telescope. It was a small spyglass, small enough to be carried about. The diameter of its lens was only 2½ inches, yet it enabled Galileo to see 100,000 stars that had never been seen before by man.

Galileo turned his telescope toward the planet Jupiter and, to his amazement,

Fig. 34-2. A painting of Galileo demonstrating his telescope from the tower of St. Mark's in Venice in 1609.

he noticed, from their positions on successive nights, that four bodies were revolving about Jupiter. He had discovered four of Jupiter's moons. This observation clearly disproved Aristotle's theory that all heavenly bodies revolved about the earth. From this and other observations Galileo concluded that the Copernican theory was correct, and he strongly supported it.

By supporting the theory that the sun, and not the earth, is the center of the universe he opposed the accepted views of his time. His writings were banned by the Church; he was brought to trial, and spent some time in prison.

THE SUN

The sun is a sphere of hot gases, 93 million miles from the earth. Its diameter (860,000 miles) is more than 100 times the diameter of the earth. From the diameters of the earth and sun it can be computed that the volume of the sun is about 1,300,000 times the volume of the earth. The movement of sunspots shows that the sun, like the earth, rotates on its axis from west to east, and that the time of a complete rotation is 25 days.

Although the sun is a gaseous sphere, we must not think that its density is the same throughout. At its surface the atoms are thinly distributed as in any hot gas at low pressure. Towards the center the gas becomes more compact so that, half way to the center, it is about as dense as water. At the center itself the density of the gas is seven times that of solid lead. It is not surprising, therefore, that the internal pressure is enormous, at least one billion tons per square inch.

The Sun's Temperature. The temperature of the surface of the sun can be computed from its brightness—it is approximately 6000° C. But the temperature of its interior is estimated to be about 20 million degrees C. We can get an idea of the temperature of the surface by remembering that the temperature of the filament of a 150-watt light bulb is about 3000° C. But probably the best way to think of the fantastically high internal temperature is to consider its devastating effects. If a coin the size of a quarter could be maintained at a temperature of 20 million degrees, the radiations from it would kill every animal and plant within a thousand miles. Only nuclear reactions can produce temperatures of this magnitude. Indeed, a temperature of this order is produced, for a fraction of a second, every time an atomic bomb is exploded.

As already stated, nuclear reactions have been going on in nature since the world began. The almost limitless energy of the sun is due to a fusion reaction in which groups of four hydrogen atoms form helium nuclei. This change is accompanied by a loss in mass and the mass is converted into energy according to the Einstein equation. Actually, in the sun 600,000,000 tons of hydrogen are transmuted to 596,000,000 tons of helium every second, so that mass is being converted into energy at the rate of about 4 million tons per second. We must not assume, however, that the sun's mass will soon be used up because, at this rate, the sun loses less than 0.01 per cent of its mass in a billion years.

Because of the high temperatures, there are no compounds in the sun's interior. Even if it were possible for, say, a molecule of water to be formed, the bonds holding the atoms together would instantly be broken. Moreover, even atoms themselves are broken. Electrons are stripped from the nuclei so that in the deep interior of the sun there are only electrons, protons, neutrons, and nuclei moving about at incredibly high speeds. Then again only the nuclei of the simple elements like hydrogen, helium, carbon, nitrogen and oxygen are present since these nuclei are the ones most easily formed.

At the sun's surface, however, where the temperature is much lower, nuclei are able to condense to form heavier nuclei. It is not surprising, therefore, that atoms such as sodium, calcium, iron, silver and even lead are present in the surface vapors. However, as might be expected, the two elements in overwhelming abundance in the sun's atmosphere are hydrogen and helium.

Light from the Sun. What is the intensity of the sun's light? This has been measured and found to be hundreds of billions of candle power. Of course, at a distance of 93 million miles the earth is a very small target, and less than one-billionth part of the solar radiation (heat and light) reaches us. Nevertheless, without this tiny fraction every living thing on the earth would perish.

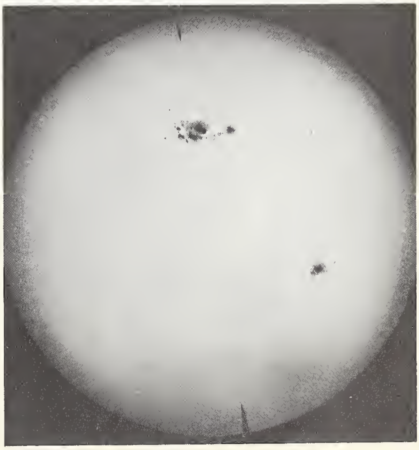
We have now referred to the temperature, pressure, and candle power of the sun. All these values are so high that they are beyond our imagination. Indeed, the word ***astronomical*** has been coined to describe magnitudes far beyond the realm of our experience.

Elements in the Sun. Scientists have analyzed the sun's atmosphere by a spectroscope. Such an analysis depends upon color absorption as indicated in the following experiment. If a beam of white light is passed through a flask containing hydrogen, and if the emergent beam is then passed through a glass prism (a spectroscope), the spectrum shows that certain colors have been absorbed by hydrogen atoms in the flask. That is to say, the spectrum is not continuous; there are gaps in it, which reveal themselves as dark lines. Hydrogen atoms always absorb light of the same wave length and therefore always give the same pattern of dark lines. The same is true for every other atom. ***Every atom gives its own particular dark-line or absorption spectrum.*** In other words, atoms (or elements) can be identified by their absorption spectra just as a man can be recognized by his fingerprints.

If light from the sun passes through a spectroscope, the spectrum is crossed by a large number of parallel dark lines. What do the dark lines indicate? As sunlight seeps up from the interior and passes through the sun's atmosphere, various atoms seize their own particular set of wave lengths from the sunlight. This loss is registered as dark lines on the spectrum. The degree of blackness of an absorption line also indicates the abundance of that particular element in the sun's atmosphere. Hence the dark lines in the sun's spectrum indicate both the number of elements and also their relative abundance in the solar atmosphere. By examining the dark lines at least 66 of the earth's elements have been found in the sun. Moreover, the relative abundance of these elements in the sun is about the same as in the crust of the earth. At least this statement is true except for hydrogen and helium, which are vastly more plentiful in the sun than in the earth.

Sunspots. From time to time dark

Fig. 34-3. A photograph of the sun showing sunspots. Notice that the spots appear on both sides of the sun's equator. (The pointers indicate the sun's axis of rotation.)

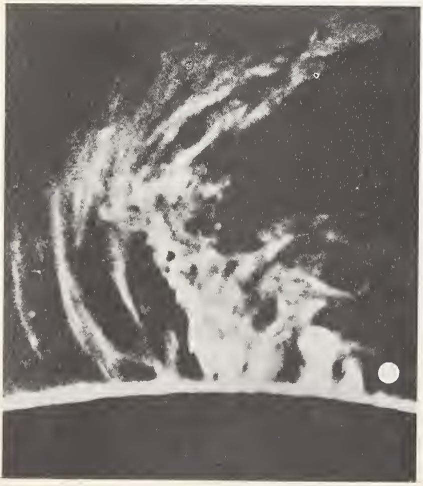


patches called ***sunspots*** appear on the surface of the sun. Sometimes the spots are more than 50,000 miles across, large enough to swallow the earth. Large ones can be readily seen by the naked eye. If you look at them, however, be sure to protect your eyes by using heavily smoked glasses. For some reason not yet understood sunspots appear on both sides of the sun's equator, but never within 45 degrees of either pole.

What are sunspots? Although sunspots cause radio black-outs on earth, astronomers are not at all sure what they are nor how they are formed. They are, however, agreed that they are the result of a deep-seated disturbance in the sun's interior. The disturbance probably causes gas to rise to the surface, where it expands and cools. It is certain that the spots look dark because they are about 1000 degrees cooler than the surrounding gases.

Solar Eruptions. Rotation of the sun carries sunspots across the surface from west to east. When a sunspot reaches the rim, gigantic flame-like prominences or curved streamers can usually be seen above the spot, sometimes rising more than 100,000 miles. Solar flares can also be seen or photographed anywhere on the sun's surface by an instrument called a spectrohelioscope. Some astronomers have compared sunspots with cyclones in the earth's atmosphere and solar flares with tornadoes.

Fig. 34-4. A solar eruption. This prominence extends more than 100,000 miles beyond the sun's surface. (The small white circle indicates the relative size of the earth.)



A solar flare is a gigantic eruption in which millions of tons of electrified particles are shot out of the sun. There are also bursts of radiations, visible and invisible, which play havoc with radio reception on earth and cause spectacular aurora displays (Northern Lights) in the skies. The electrons and protons must travel at speeds of more than 1000 miles a second, because they affect sensitive magnetic instruments in observatories all over the earth on the day following a solar eruption. They travel towards the poles of the earth and bombard air molecules at heights of 50 to 100 miles, tearing electrons out of oxygen and nitrogen atoms. This disruption of atoms causes them to give off light by a process similar to the operation of a neon light.

The ejection of electrons, however, does not explain the blacking-out of short-wave radio transmission which occurs almost as soon as a flare is observed. This effect is probably due to a blast of ultraviolet light and X rays which, you will recall, can travel from sun to earth in 8 minutes or so. When this short-wave radiation strikes the ionosphere it ionizes a great many air molecules, that is, it knocks off some of their electrons, thereby leaving the atoms electrically charged. The result is that this region of the atmosphere is then unable to reflect radio waves back to earth. Instead, the radio waves escape into outer space and a black-out results.

The Sun Gives Life to the Earth. In earlier chapters we have seen how important the sun is to us. Without the light and heat of the sun there could be no life on the earth. The most vital of all chemical reactions is photosynthesis,

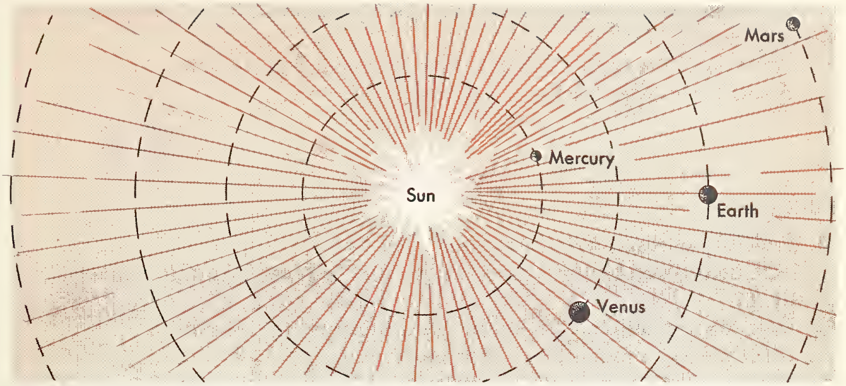


Fig. 34-5. Only a minute fraction of the sun's radiation strikes the earth. Yet, without this minute fraction, there would be no life on earth.

which depends on light. Without sunlight plants cannot make chlorophyll; and without chlorophyll they cannot make starch and sugar, which are the main foods of both plants and animals. In other words, if there were no sunlight, there would be no vegetation, and without vegetation animal life could not exist on the land or in the sea. Sea life, both plant and animal, is the source of petroleum; thus without sunlight we would have no natural sources of gasoline.

The production of food and petroleum are not the only important results of photosynthesis. All plant cells have a membrane of cellulose. This is also true of trees, where the cellulose forms the hardy woody structure. Without sunlight there would be no cellulose; without cellulose there would be no trees; without trees there would be no coal; without coal and oil our modern civilization could not have developed.

Thus we see that life on the earth depends on the ***light*** of the sun; but the ***heat*** of the sun is as necessary to life as the light. Without the sun's heat there would be no evaporation from rivers, lakes, and seas. Without evaporation there would be no rainfall, and without rain there would be no water for the crops of the earth. Indeed, without the sun the temperature of the whole earth would be so low that all rivers, seas, and lakes would be solid ice.

If the earth's surface were not heated by the sun, there would be no winds, for winds are convection currents formed by the unequal heating of the earth's surface. Without winds there would be no circulation of water vapor, which is another essential factor in sustaining life on earth.

Things to Remember

Copernicus, a Polish astronomer, stated in the 16th century that the sun, not the earth, is the center of a planetary system.

Galileo made the first telescopic observation of the heavens. He discovered four of Jupiter's moons.

The ***sun*** is an enormous mass of gas whose temperature and pressure increase towards the center.

The ***temperature of the sun*** at the surface is about 6000° C; at the center it is about 20 million° C.

At least 66 of the earth's elements have been found in the sun.

Sunspots cause radio black-outs on earth.

Questions

GROUP A

1. What are obelisks? What were they used for?
2. "Aristarchus was centuries ahead of his time." What does this statement refer to?
3. State the meanings of the words, (*a*) astronomy, (*b*) planet.
4. What are (*a*) sunspots, (*b*) solar prominences?
5. What are the approximate temperatures of the sun (*a*) at its surface, (*b*) at its center?

GROUP B

6. What were the contributions of (*a*) Copernicus, (*b*) Galileo, to astronomy?
7. What can you say of the nature of matter in the sun's interior? What is the evidence to support your ideas?
8. (*a*) What is the spectroscopic evidence of elements in the sun's atmosphere? (*b*) How does a spectroscope reveal definite elements?
9. Explain the connection between sunspots and (*a*) Northern Lights, (*b*) radio black-out.
10. "All life on earth depends on the sun." Justify this statement.

Things to Do

Viewing sunspots · Sunspots can be "seen" by using a small telescope. On no account, however, should you look directly into the sun—the sunlight would probably blind you. Instead, sunspots must be observed indirectly.

Direct a small telescope towards the sun (a supported pocket telescope is satisfactory). The shadow cast by the telescope will tell whether or not the telescope is aligned correctly. Pick up the image of the sun on a white card held a few inches from the eyepiece, and move the card back and forth until the sun is in sharp focus. Sunspots (if they are present) can now be seen as black dots on the card.

35

THE SOLAR SYSTEM

FROM EARLIEST TIMES man has studied the heavens. He noticed that some "stars" appeared to move and that others seemed to be fixed. Then, about 300 years ago, came the telescope, a device that vastly increased the number of visible stars. The telescope also picked out additional "movable stars," and today we know there are at least nine of them.

"Movable stars" are really not stars at all; they are planets. What is the difference between stars and planets? Due to nuclear activity stars, like the sun, emit their own light and heat. Nuclear activity does not take place in the interiors of planets, so they shine only because they reflect the light of the sun.

All nine planets revolve in the same direction about the sun in slightly elliptical (that is, almost circular) orbits. Moreover they revolve very nearly in a common plane. In this respect, the solar system differs from an atom with which it has been compared.

The six planets nearest the sun—Mercury, Venus, Earth, Mars, Jupiter, and Saturn—are the brightest and have been known since earliest times. Because of their great distance from the sun, the three outermost planets are faint and were not discovered until comparatively recent times; Uranus in 1781, Neptune in 1846, and Pluto in 1930.

The time taken for each planet to travel around the sun is called its ***year.*** A year is, of course, different for each planet: Mercury's year is equal to 88 of our days, whereas Pluto's year is equal to 250 of our years.

Knowing the radii of their orbits and the lengths of their years, it is fairly easy to compute the speeds of the planets around the sun. Let us consider the speed of the earth. We know that the distance from earth to sun is 93 million miles. This is the radius of the earth's orbit which we shall assume to be circular. Hence the circumference of the orbit ($2\pi \times$ radius) can easily be computed. This is $2 \times 3.14 \times 93{,}000{,}000$ or 584,000,000 miles.

We also know that the earth takes one year or 365¼ days to complete its orbit. The number of hours in one year is 8760 ($365\frac{1}{4} \times 24$). Hence the velocity of the earth's revolution around the sun is

$$\frac{584{,}000{,}000}{8760} = 66{,}600 \text{ miles per hour}$$

or 18.5 miles per second.

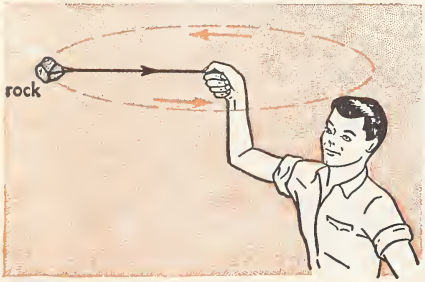
Why Do Planets Move in Curved Paths?

Although early astronomers made many observations on planets and stars, the "laws of the heavens" were a complete mystery to them. It was not until the latter part of the 17th century that Isaac Newton, the famous English physicist, discovered the Law of Gravitation, a law which explains the forces between all heavenly bodies. The law states: ***Any two bodies attract each other. The greater the mass of the bodies, the greater is the force of attraction between them; the greater the distance between the bodies, the less is the attraction between them.***

Newton is supposed to have discovered this universal law when he asked himself the question, Why does an apple fall from a tree to the ground? According to the law, the earth attracts the apple with the same force that the apple attracts the earth. This force moves the smaller apple much more easily than it moves the larger earth and, as a result, we see only the motion of the apple. Similarly, if we consider the earth and sun as a pair we can say that the sun attracts the earth just as the earth attracts the sun. However, the earth is so much smaller than the sun that its motion is more affected by the sun's attraction than is the sun by the earth's attraction. (We have already considered tides as a gravitational effect on page 16.)

You can demonstrate the motion of a planet around the sun by attaching a rock to the end of a string and whirling it above your head. In whirling the rock in a circular path you realize that your hand is exerting an inward force on the string at right angles to the path of the rock. (See Fig. 35-1.) This force pulls the rock from a straight line path into a circular one; it corresponds to the gravitational attraction of the sun on the earth. If the string were suddenly released the rock would fly off at a tangent to the circle.

Fig. 35-1. The inward pull of the hand on the rock corresponds to the gravitational pull of the sun on the earth. The earth, like the rock, moves in a circular path.



The earth moves in a similar way. Its curved path is a compromise between the direction due to the law of motion and the direction due to the gravitational force. Based on the gravitational law, astronomers can predict the exact path of every planet and comet around the sun.

Origin of the Solar System. Actually nothing is known definitely concerning the origin of the solar system. During the 18th and 19th centuries a number of theories were put forward by astronomers and mathematicians, but all had some serious defect. Most of the older theories, for example, assumed that the planets were originally a part of the sun. There are several objections to this theory. Consider for example, the composition of the sun and the earth. The earth has a high percentage of heavy elements, such as silicon 25%, aluminum 7%, iron 4%; the sun, on the other hand, consists predominantly of hydrogen and helium. Then again, both Saturn and Jupiter rotate 60 times faster than the sun. It is difficult to see how this could be so if they were once part of the sun.

Probably the most satisfactory theory is a relatively recent one known as the ***dust-cloud theory***. It is known that there are vast collections of gas and "dust" in space. Dust is made up of microscopic particles and some astronomers believe that the mass of suspended dust is as great as that of all the stars together. Gravitational forces pull the particles closer together so that, eventually, a large segment of cloud begins to contract. As particles gravitate towards the center the temperature rises and they begin to rotate. As the cloud condenses more and more, its temperature gets higher and higher. After it has shrunk in volume by a billion times or so the temperature and density are high enough for collisions between hydrogen atoms to produce large-scale nuclear reactions in the star (or sun). When nuclear reactions begin, the

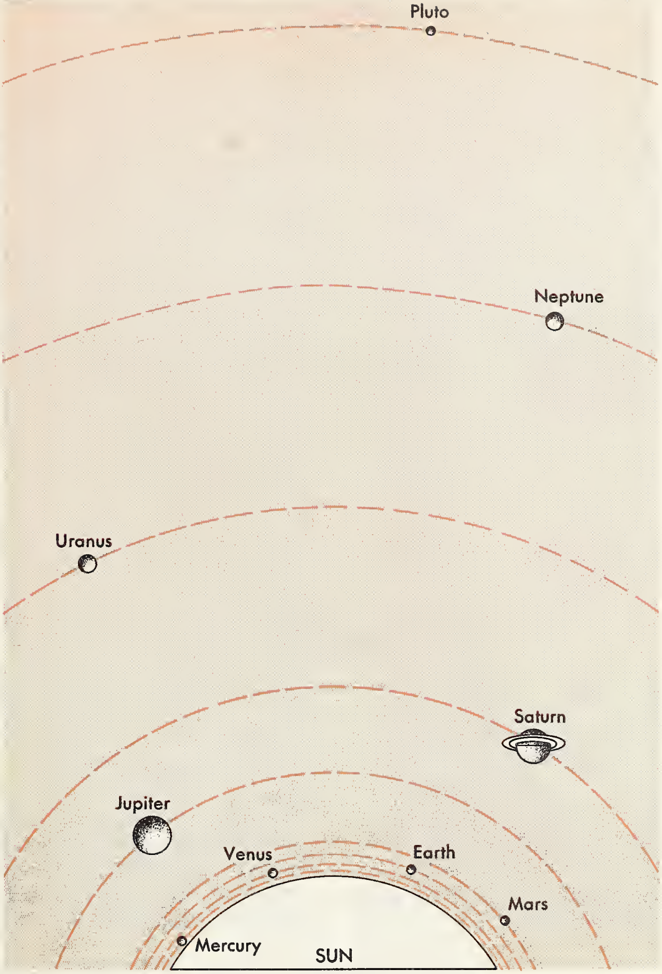


Fig. 35-2. The inner planets are Mercury, Venus, Earth, and Mars; the outer planets are Jupiter, Saturn, Uranus, Neptune, and Pluto.

star becomes stable; that is to say, it no longer contracts in size.

How were the planets formed? When the main cloud collapsed to form the sun, a number of spiraling masses were left stranded out in space. These condensed to form the planets. The heat of contraction made the planets very hot, probably hot enough to melt them. But all the planets are much smaller than the sun, so that their heat of contraction was far less than the sun's. As a result, the heat in the interior of planets was not enough to start nuclear reactions. Instead they slowly cooled off, and the cooling process continues to this day.

THE PLANETS

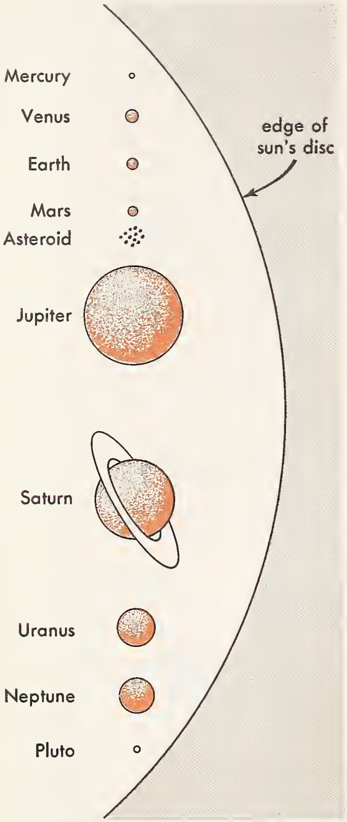


Fig. 35-3. The approximate sizes of the planets compared with the sun.

The planets do not revolve about the sun at the same speed. The nearer a planet is to the sun, the greater is its gravitational attraction. Moreover, the greater the force of attraction, the greater must be the speed of a planet. The speeds of the nearer planets must be exceedingly high, high enough to counteract the large gravitational force. Otherwise, they would be pulled into the sun.

The planets can be conveniently divided into three groups: (1) those nearest the sun—Mercury, Venus, Earth, and Mars; (2) the asteroids; and (3) the outer group—Jupiter, Saturn, Uranus, Neptune, and Pluto. The mass of all the planets together is not more than 0.2 per cent of the mass of the sun. The planets near the sun are the smallest (the dwarf planets). Moreover, their distances from the sun and from each other are relatively short. The outer ones (except Pluto) are giant planets, and their distances from the sun and from each other are vast. Jupiter is 5 times as far from the sun as is the earth; Saturn is 10 times as far from the sun as is the earth; Uranus 20 times; Neptune 30 times; Pluto 40 times.

Between the inner and outer planets (actually between Mars and Jupiter) is a broad rotating girdle made up of pieces of rock of various sizes called asteroids. They are discussed on page 363.

THE INNER PLANETS

Mercury. Mercury is the smallest of the planets (it is not much larger than our moon), and the nearest to the sun. The name ***Mercury*** was given to this planet by the ancient astronomers because it always appears in the sky close to the sun—Mercury, in Greek legend, was a close friend of Apollo, the sun god.

There is a strong gravitational force between the sun and Mercury. Why? What are some of the effects of this powerful force of attraction? First, Mercury is held in the grip of the sun so that the same half always faces the sun. According to the dust-cloud theory, Mercury must have rotated rapidly when first formed, but the sun has braked its rotation until it now only rotates on its axis once each time it revolves about the sun.

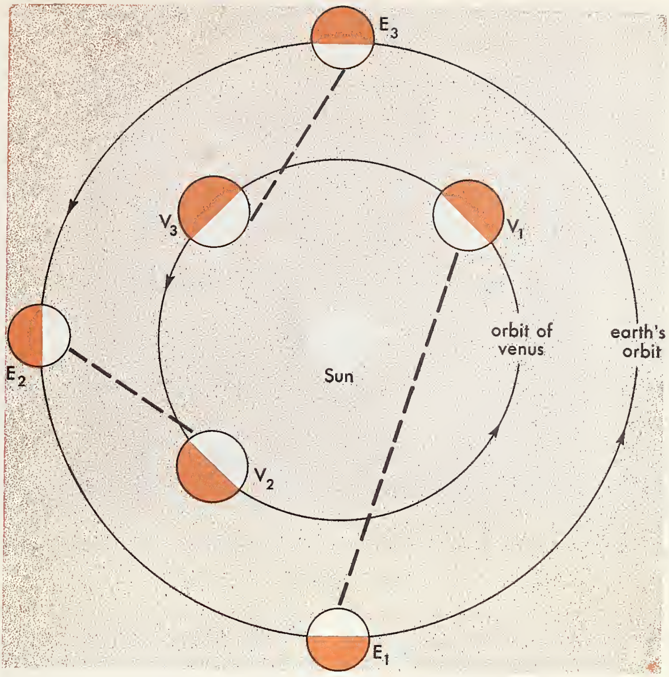


Fig. 35-4. Venus, like the moon, shows phases.

What of the temperatures of Mercury? The side facing the sun is in perpetual sunshine and is blisteringly hot—it is about as hot as molten lead. The other side is in everlasting darkness and exceedingly cold—about as cold as dry ice.

And, finally, what of its atmosphere? Mercury has no atmosphere. It probably had one originally, but it was too small to hold the gaseous envelope. That is to say, because of its small size its gravitational force is so small that the gas molecules in its atmophere probably escaped into space.

Venus. Venus is about the same size as the earth and, except for the sun and moon, is the most brilliant object in the sky. Indeed, at its greatest brilliance, it is sometimes plainly visible during the day. Venus, like Mercury, revolves in a smaller circle than the earth. For this reason it is always seen near the sun, and most easily seen in the early morning or early evening. It is often called the Morning or Evening Star.

Because it is between the earth and the sun, Venus passes through phases like those of the moon (see Fig. 35-4). Venus (V_1) is seen as a "full moon" from the earth (E_1); V_2 is seen as a "half moon" from E_2 and V_3 is seen as a "crescent" from E_3.

The view of Venus, even through a large telescope, is disappointing. No physical features can be seen because the planet is covered by a dense cloud layer. The spectroscope suggests that the atmosphere of Venus contains little or no oxygen but a great deal of carbon dioxide, more than 400 times as much as in the earth's atmosphere. Because of the absence of oxygen, it is generally believed that there is no life on Venus.

Mars. Mars is our nearest neighbor,

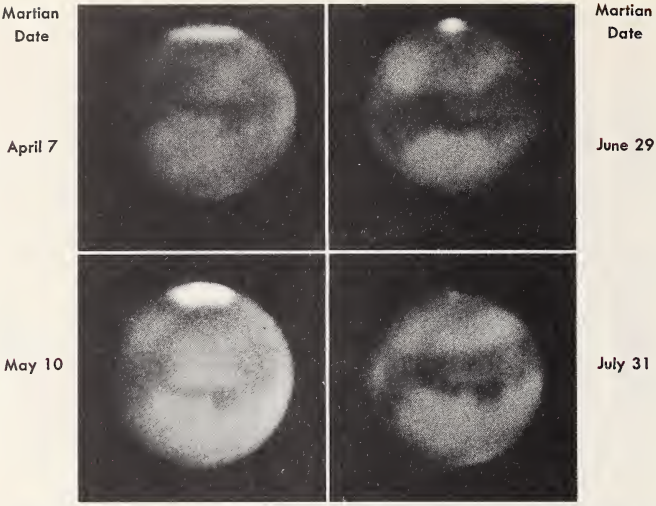


Fig. 35-5. This series of photographs shows seasonal changes on Mars. Notice the shrinking of the polar icecap and the increase in the size of the dark area (the green belt) as the summer advances.

40 to 50 million miles away. Its brilliance (almost equal to that of Venus) and its red color make it a familiar object in the sky. The diameter of Mars is about half that of the earth. The length of its day is about the same as ours, but it takes almost two years to travel round its orbit. From the above facts, what conclusions can we make concerning the conditions on Mars?

First, there is far less heat and light on Mars than on the earth. Why? The temperature of Mars has been carefully studied. Telescopes concentrate both light and heat rays. Sensitive instruments attached to telescopes enable astronomers to measure the temperature of planets with considerable accuracy. We know the temperature of Mars is far lower than that of the earth. The mean temperature of the earth is about 60° F, that of Mars is −30° F. Even in the tropics the temperature of Mars does not exceed 70° F. And because of its thin atmosphere (atmospheric pressure at the surface of Mars is only 65 mm.) the heat of the day is rapidly lost so that the nights are very cold, probably −70° F even in the tropics.

What of the atmosphere on Mars? Because Mars is small its gravitational hold on the atmosphere is weak and, as already stated, its atmospheric pressure is about one-tenth of the earth's pressure. There is very little oxygen or water vapor in its atmosphere; carbon dioxide is about twice as plentiful as on the earth, and the main ingredient is nitrogen.

The most conspicuous features on Mars are the polar caps which are easily seen through a telescope. They advance in winter and retreat in summer and are never entirely absent from the Martian scene. Since there is very little water on Mars, these caps are certainly not deposits of deep snow. More probably they are areas of thin ice and frost.

Photographs of Mars, particularly those taken in the spring, show a number of straight lines on the surface. For more than a century some astronomers have argued that the lines may be "canals" which carry water from the polar regions to irrigate the dry interior. However, canals would have to be 10 miles wide to

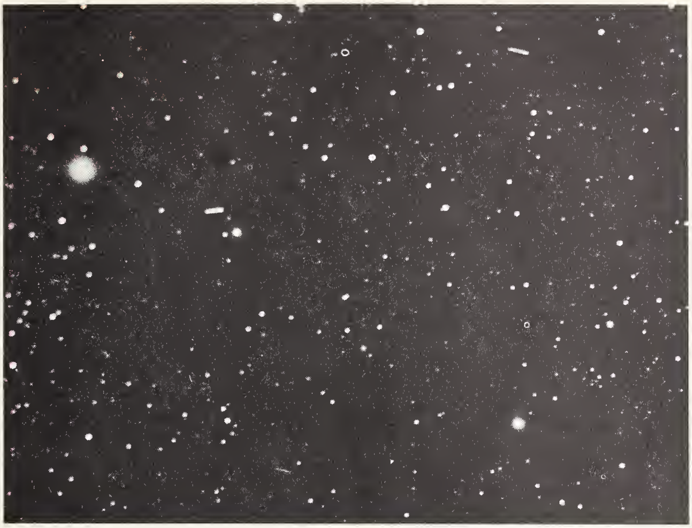


Fig. 35-6. Asteroid tracks. To make this photograph, an exposure of about an hour was made with a telescope-camera. The trails of three fast-moving asteroids can be seen in this picture.

be observed at a distance of 50 million miles. And so the regular markings remain a mystery.

How can we account for the red colors of Mars? It is thought that minerals have combined with and used up most of the atmospheric oxygen. These reddish mineral products now form a vast sandy desert that covers almost three-quarters of the surface of Mars. Patches of green, perhaps plant life, can also be seen.

Is there life on Mars? Are there such people as Martians? Certainly the construction of canals assumes intelligent beings, but do they really exist? The absence of oxygen, a shortage of water, and a forbidding climate are conditions that preclude animal life as we know it on earth. But what of plant life? The necessary ingredients, water vapor and carbon dioxide, are present in the atmosphere. It is generally conceded that mosses, lichens and algae, lower forms of plant life, may survive the rigors of the Martian climate.

The Asteroids. In the 19th century astronomers and mathematicians had shown by their computations that there ought to be a planet between the inner dwarfs and the outer giants. Actually a tiny "planet," about the size of England, was picked up by a telescope in the computed orbit. But when photography was introduced into astronomical research a large number of tiny planets, almost 2000 of them, were spotted. In practice, photographic plates were exposed for several hours, and when they were developed the paths of asteroids showed up as tracks or lines on the plates. (See Fig. 35-6.)

We now know that the asteroids are chunks of rock that vary in size from a few miles in diameter to 400 miles or so. As one might expect, they all travel in the same direction. But the surprising fact is that all the asteroids together would comprise a body no larger than the moon. How can we account for their origin? Here again is a good deal of uncertainty. Some astronomers have suggested that a small planet traveling in this orbit may have been torn apart by gravitational stresses. The nearby giant Jupiter, with its large gravitational attraction may have drawn many of the fragments into its orbit. Similarly, Mars, on the other side, may also have captured some. The residue, the uncaptured ones, constitute the asteroids.

THE OUTER PLANETS

Jupiter. Jupiter is the largest of the planets, larger than all other planets together. It is about 1000 times the volume of the earth. Its period of rotation is only 10 hours. This unusually rapid rotation (which is 60 times as fast as the sun's) is one of the reasons why some astronomers believe that Jupiter was never a part of the sun.

Jupiter has 12 satellites or moons. The four largest were the first heavenly bodies to be discovered by a telescope. You will recall that Galileo first observed them in the 17th century, and that this observation convinced him of the validity of the Copernican theory.

Saturn. Saturn, the second largest planet, is 750 times larger than the earth. It is the most distant of the planets visible to the naked eye. Like Jupiter it rotates rapidly and, also like Jupiter, the solid planet is hidden by clouds of ammonia.

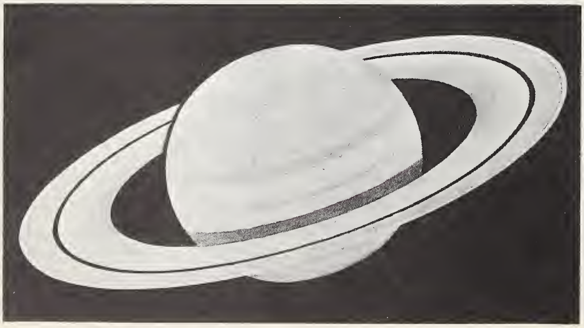
Saturn has no fewer than eight moons, but its most spectacular feature is its three flat circular rings which rotate about the planet. Actually, it is one broad ring, divided into three parts; it looks like the brim of a straw hat. The rings are composed of millions of small solid bodies, which are probably the remains of moons that got too near the planet and disintegrated.

Again Galileo was the first astronomer to observe Saturn's rings. A telescope, however, is unnecessary. Both the rings and the four moons of Jupiter can be seen with a pair of field glasses.

Uranus. Uranus is about half the size of Saturn. It is barely visible to the unaided eye and, for this reason, was never detected by the ancient astronomers. It was discovered in 1781 by the famous English astronomer, William Herschel. Herschel, scanning the heavens with a telescope, happened to see a faint disc which he continued to observe for several months. The disc moved among the stars and it was apparent that he had discovered a new planet. From his observations he computed its orbit and named it Uranus, from a Greek work meaning ***heaven***. It is interesting to note that the metal uranium was discovered by Klaproth the same year that Herschel discovered Uranus. The new metal was named in honor of the new planet.

Neptune. Neptune, about the same size

Fig. 35-7. Saturn is encircled by three flat rings that look like the brim of a straw hat.



as Uranus, is invisible to the naked eye. It was discovered in a most interesting way. Discovery of the planet by direct observation would have been exceedingly difficult because it is so dim. Instead, it was discovered indirectly with the aid of Newton's law of gravitation.

Newton's laws of motion enable astronomers to predict the exact time a planet will be at a definite place. Any irregularity in its motion indicates that some other planet is disturbing it. Observations showed that Uranus was being deflected from its calculated orbit, and Newton's law indicated that a planet beyond Uranus might account for the disturbance. In 1846, two astronomers, Leverrier of France and Adams of England, independently determined the position of the disturbing body. Leverrier sent the information to Galle at the Berlin Observatory. Galle, in turn, turned his telescope to this region of the sky and, almost at the exact spot predicted, the new planet Neptune was discovered. The discovery of Neptune came a century and a half after the death of Newton. It was a remarkable confirmation of the universal nature of the gravitational law.

Pluto. Pluto is much smaller than Uranus or Neptune. Because it is so small and so distant from the sun it is difficult to see, even with the best telescopes. Pluto, like Neptune, was also discovered indirectly. It was found that Neptune was deflected from its calculated orbit, and this led astronomers to believe there was still another planet beyond Neptune. Percival Lowell at Flagstaff, Arizona, made his calculations and scanned the heavens at the selected site. Lowell died before the search was completed but, in 1930, Pluto was discovered by photographs in the region predicted by Lowell. Pluto was named after the god of darkness because it receives so little light from the sun. Strangely enough, the first two letters of Pluto are also the initials of Percival Lowell.

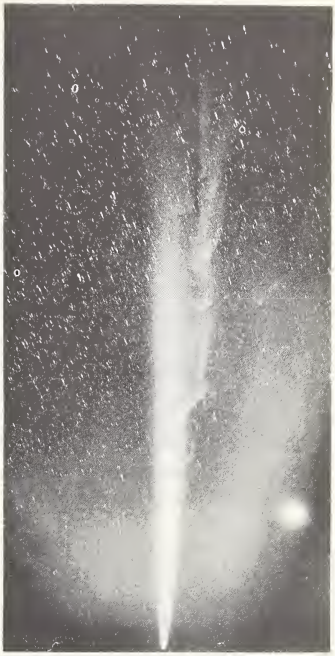


Fig. 35-8. Halley's Comet, photographed during its last appearance in 1910. The bright circular object on the right of the photograph is the planet Venus.

Are there any planets beyond Pluto? This question cannot, of course, be definitely answered. Large planets like Jupiter or Uranus would deflect Pluto from its orbit so that they could be discovered. But smaller ones might escape detection for years.

COMETS AND METEORS

Comets. Comets are the most spectacular members of the solar system. Large comets, with their long tails sweeping across the night sky, can be seen by the naked eye. To the ancient observers they were an ill omen—a sign of impending disaster.

What are comets? Comets are composed of gases, dust particles, and small solid particles as large as peas, all held to-

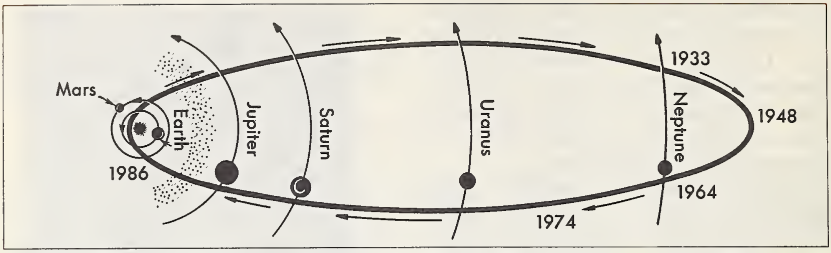


Fig. 35-9. Halley's Comet covers its elliptical path every 76 years. Its next appearance will be in 1986.

gether by gravity. As a comet approaches the sun, the sun's radiation exerts enough pressure to force the gases away from heavier particles to form a long tail. A comet therefore consists of a nucleus and a tail, the tail always pointing away from the sun. Comets, like other members of the solar system, revolve about the sun in elliptical orbits.

The best-known comet is Halley's Comet, named after an English astronomer who made observations on it in 1682. Halley calculated its orbit and predicted its return in 1757. This was the first prediction of a comet's return. Actually Halley's Comet appears regularly every 76 years. Its last appearance was in 1910, so that it will not again be visible until 1986. The nucleus of Halley's Comet is not more than 20 miles in diameter, but its tail is millions of miles long.

Fig. 35-10. The trail of a meteor. The brilliance of the meteor shows up by comparing it with the dimmer star cluster in the background.

How are comets formed? The origin of comets is unknown, but various theories have been suggested to account for them. One is that, millions of years ago, the sun captured the particles from an interstellar cloud that came under its gravitational influence.

Meteors and Meteorites. If the earth passes through the orbit of a comet we may observe a shower of meteors (sometimes called "shooting stars"). Meteors are probably remnants of comets. Most meteors are about the size of a pinhead and, when they reach the earth's atmosphere, they travel at about 30 miles a second. At 40 to 60 miles above the earth the air is dense enough to heat meteors to incandescence by friction, and they burn up. Astronomers tell us that at least a billion meteors reach the earth's atmosphere every day.

If a meteor is large enough to survive the trip and reaches the earth it is called a ***meteorite.*** Most meteorites are stony; some are iron; a few contain nickel. The largest meteorite to fall in the United States landed in Arizona some 50,000 years ago. It formed a crater nearly a mile wide and 600 feet deep. Computations show that to carve this enormous crater the meteorite must have weighed at least 10 million tons and it must have struck the earth with a speed of almost 150,000 miles an hour (about 40 miles per second).

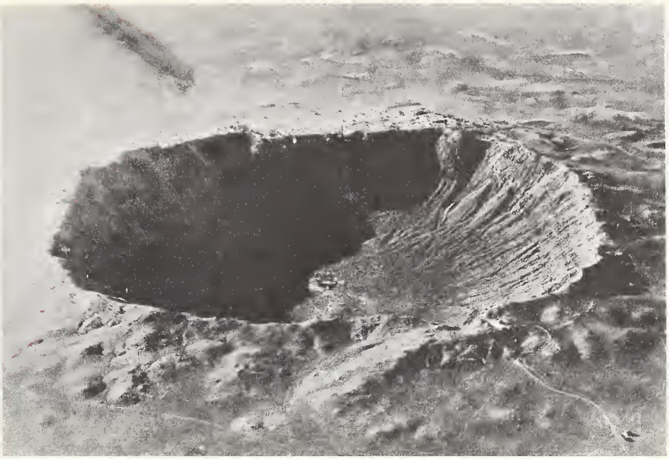
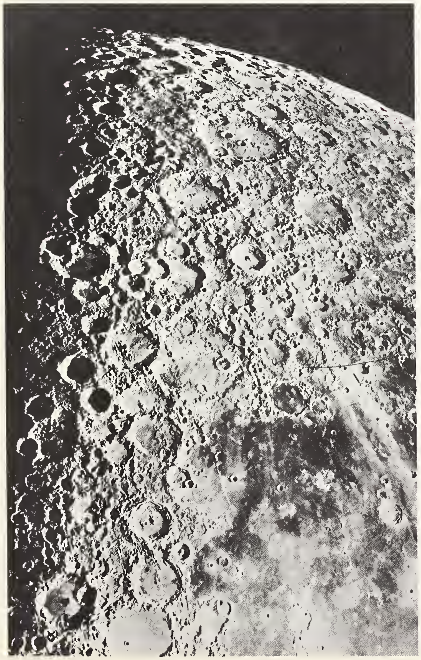


Fig. 35-11. Meteorite Crater, Arizona. Millions of tons of rocks were displaced when an enormous meteorite, probably a small asteroid, fell to earth in Arizona some 50,000 years ago.

THE MOON

Fig. 35-12. The face of the moon is covered with craters.



The moon is a comparatively small body. It is only 2000 miles in diameter, about $\frac{1}{60}$ of the size of the earth, yet it shines so brilliantly and seems so near (it is only a quarter of a million miles away) that it has attracted man's attention since the dawn of history. Indeed, the moon can be seen better with the naked eye than any planet can be seen through a telescope. Astronomers, since the time of Galileo, have drawn maps showing physical features of the moon.

There is no water on the moon—just light and dark areas of land. The dark areas are great plains. Galileo mistook them for oceans and called them "maria" (the Latin word *mare* means "sea"). The term is still used to this day. There are mountains on the moon, some of them higher than Everest. But the most interesting features are its craters. The moon's face is literally covered with craters, some far deeper (four or five times deeper) than the Grand Canyon. What caused the craters? No one knows. The best guess is that they are either extinct volcanoes or craters left by large meteorites.

The moon is held in the gravitational

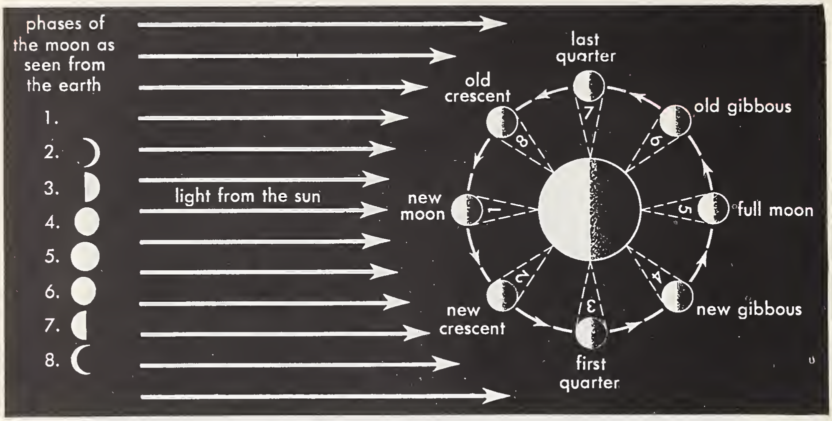


Fig. 35-13. Phases of the moon. The diagrams on the left (numbered 1 to 8) show what the moon looks like from the earth at each of the corresponding phases.

grip of the earth and therefore always keeps the same face towards the earth. That is to say, it rotates once on its axis as it revolves around the earth. In rotating it completes the cycle of changes from new moon to full moon to new moon. The time occupied by these changes is called a ***month,*** which is approximately 29½ days.

Like the earth, the moon is lighted by sunlight which it reflects. As the moon travels round the earth, varying amounts of it become visible. These are called ***phases*** (see Fig. 35-13). First we see a new moon (a thin crescent) in the west. As the moon travels eastward along its orbit more and more of the moon becomes visible until, finally, it appears as a disc. This is full moon. As the moon continues along its orbit less and less of it becomes visible until finally it becomes invisible.

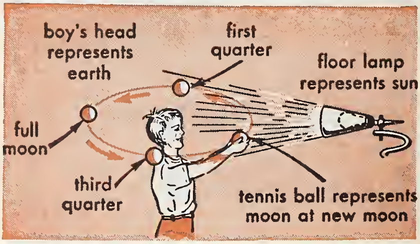
Conditions on the Moon's Surface. If we could travel to the moon by rocket what would we find? There is no atmosphere on the moon—gravitational attraction is much too small to retain an atmosphere.

Since there is no atmosphere there is no wind. Nor is there any moisture. Hence, there are no clouds, no rain, or snow, or ice. Consequently there has been no erosion through the ages—little to wear the jagged edges of its mountain peaks.

From the moon we would see the sun surrounded by stars in a black sky. Why? Stars would also appear brighter than when seen from the earth. How do we know this? But we would see no beautiful sunsets. Why not?

Daylight on the moon would last for two weeks and would be followed by two weeks of darkness. How do we know? During the day the temperature would rise to above 200° F and during the night it would drop to about −240° F. Why are these temperatures so much more extreme than on the earth?

Fig. 35-14. Why the moon appears to change shape.


boy's head represents earth
first quarter
floor lamp represents sun
full moon
third quarter
tennis ball represents moon at new moon

Nor would there be any sounds on the moon. Not even the crash of a large meteorite could be heard. Why not? Indeed, we would find the moon to be a grim, frightening but fascinating place—a dead and silent world.

Things to Remember

There are nine planets in the ***sun's family.***

All the planets ***shine by reflecting*** the light of the sun.

Mercury is the smallest planet and the one nearest the sun.

Venus is the most brilliant planet in the sky—brighter than any star.

Mars is the only planet (apart from the earth) that could support life.

Jupiter is the largest planet.

Pluto is the most distant planet.

Comets are masses of gas, dust, and small particles that travel around the sun in elliptical orbits.

Meteors are remnants of comets.

Meteorites are remnants of meteors that fall on the earth.

Phases of the moon are caused by the revolution of the moon around the sun.

Questions

GROUP A

1. Name in order the planets (*a*) between the earth and sun, and (*b*) beyond the earth.
2. Mars' year is 1.88 of our years. What does this mean?
3. The planet Mercury has no atmosphere. Why not?
4. What are the asteroids? Where are they?
5. (*a*) How did the planet Uranus get its name? (*b*) How did the element uranium get its name?
6. What is a comet? Give an example.
7. What is the difference between a meteor and a meteorite?

GROUP B

8. Mercury is 36 million miles from the sun and Mercury's year is 88 days. Compute the speed of the planet.
9. How does the dust-cloud theory account for the formation of the sun and the earth?
10. Write a brief description of the planet Venus.
11. State some of the arguments for and against the possibility of life on Mars.
12. Explain the circumstances that led to the discovery of (*a*) Uranus, (*b*) Neptune.
13. (*a*) How are comets formed? (*b*) Write a brief account of Halley's comet.
14. (*a*) Describe some of the physical features of the moon we should see if we traveled there by rocket. (*b*) Explain why no one has ever seen the other side of the moon directly from the earth.

Things to Do

Phases of the moon · Stand in a darkened room facing a source of light (a shaded light bulb is satisfactory). Hold a ball towards the light at arm's length (see Fig. 35-14). The light bulb represents the sun, the ball represents the moon, and your head represents the earth. Notice, in this position, that the section of the ball you see is in darkness.

Now, still holding the ball at arm's length, start turning your body counterclockwise. First the moon is crescent-shaped, and when you have made a quarter of a turn (one right angle) one-half of the moon's face is illuminated. This is the first quarter. When you have made a half turn (two right angles) the whole face of the moon is illuminated. This is full moon.

Continue the turning movement. How much of the face of the moon do you see after three-quarters of a turn? Notice particularly the shape of the waning moon near the completion of the turn.

36

STARS AND GALAXIES

Man's exploration of space has advanced in three steps or phases. By the end of the first phase he had a picture of the solar system—a central sun and its revolving planets. The idea of a solar system, you will recall, began with Copernicus. The details of the system were not filled in until three centuries later.

During the past three centuries telescopes have been improved and cameras invented. With their aid the concept of a ***stellar system*** began to emerge. The sun was found to be merely one of millions of suns (or stars) that constitute our own particular system, which is called the Milky Way. The picture at the end of the second phase was that the Milky Way is a vast swarm of stars that drifts through space, just as a swarm of bees flies through air.

Today astronomers explore the depths of space with a 200-inch telescope that has a light-gathering power of a million eyes. Far beyond the boundaries of the Milky Way they find millions of stellar systems similar to our own. But these systems are so far away that it is almost impossible to distinguish individual stars. For the most part they appear as patches of light and, for this reason, they are called "nebulae," a Latin word meaning ***clouds.*** Thus by the end of the third phase exploration has extended to outer space and, with it, has come the modern concept of the universe.

STAR DISTANCES

The distances of stars are truly astronomical. Indeed, a special unit, a ***light-year,*** is used to measure these distances. ***A light-year is the distance light travels in one year.*** You will recall that light travels at a speed of 186,000 miles per second. The actual number of miles in a light year is, therefore, 186,000 × 60 × 60 × 24 × 365¼, which is approximately 6,000,000,000,000 (six trillion) miles. Light from the moon reaches the earth in 1.3 seconds; from sun to earth takes about 8 minutes; from sun to Pluto about 5 hours; from the nearest star (Alpha Centauri) it takes about four and one-half years; from the nearest nebula (Andromeda) 2 million years; and from the most distant nebulae as much as two billion years. This scale is so immense that it may be well to get a clearer concept of star distances.

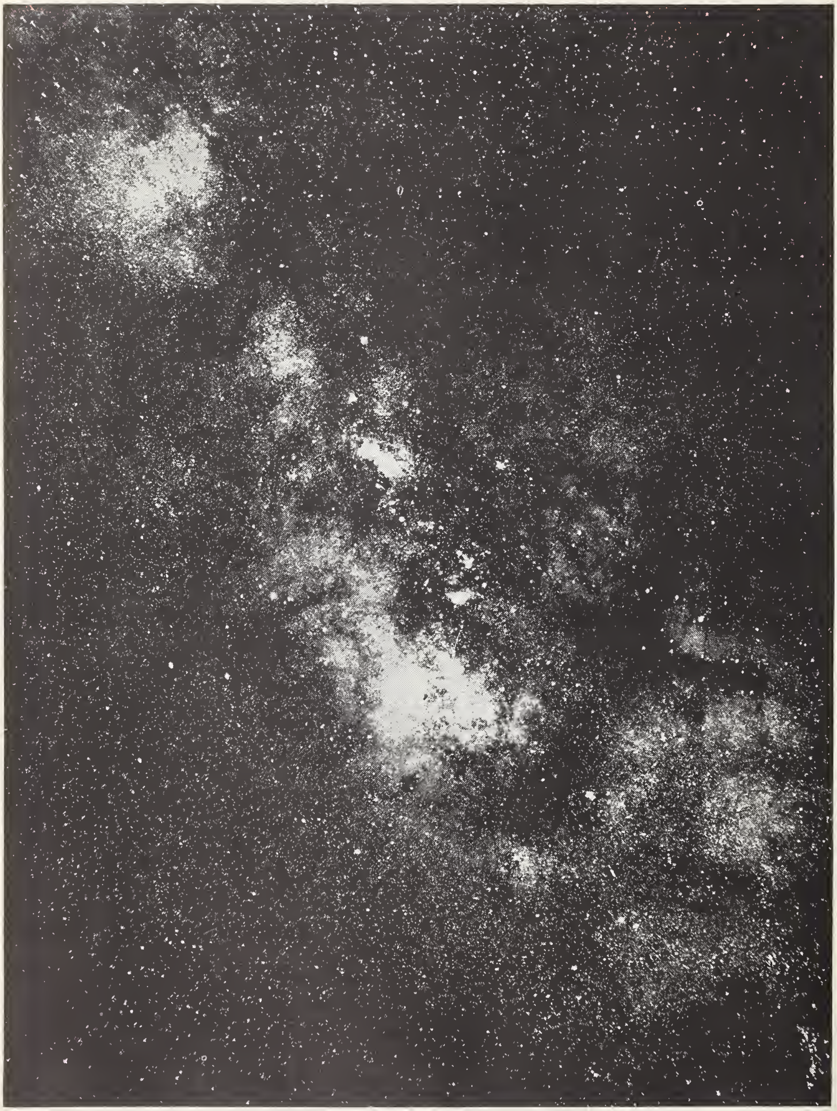


Fig. 36-1. View towards the center of the Milky Way where the concentration of stars is greatest.

36-1. Comparison of planet and star distances

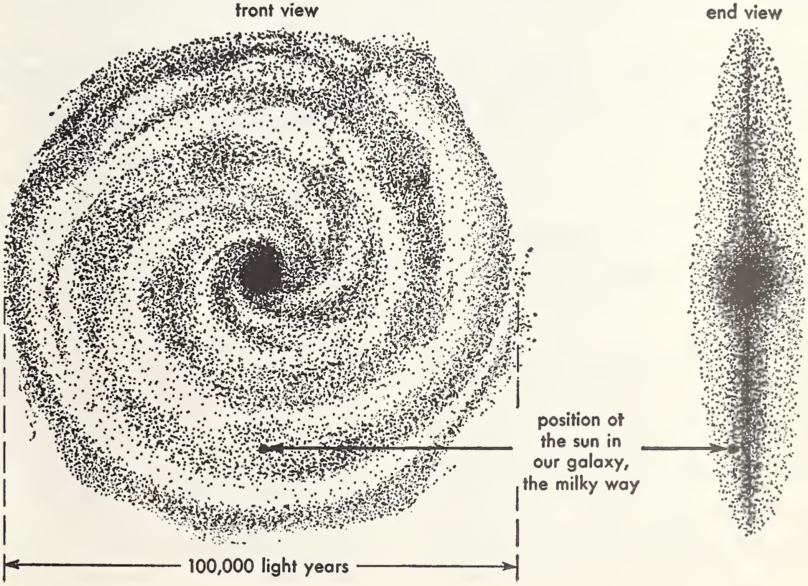
Take a large roll of paper tape like that used in an adding machine. Mark a starting point for the position of the sun. Taking a unit of 1 millimeter for the distance traveled by light in one minute, locate the position of the earth and the position of Pluto on the roll. (The distance from the sun to Pluto is 3,700 million miles, and from the sun to earth 93 million miles.) Now calculate the number of minutes in 4.5 years, and place a mark on the roll for the position of the nearest star. Explain your results in terms of distance.

You may have noticed a luminous band in the sky on a clear night, a band that stretches from the North Star to the southern horizon. Indeed, from observations made below the equator we know that this pearly band circles the earth. If you examine the band with a pair of binoculars you will find it is composed of countless stars, 100 billion of them. From its whitish appearance the ancient Greeks called it the Milky Way. Or rather, they called it the Galaxy, which is the same thing. (The Greek word ***gala*** means ***milk***.)

All the stars we see with the naked eye, and by far the largest number we see through a powerful telescope, are in this Milky Way system. The sun is just an ordinary star in the system. Some stars are larger than the sun, others smaller; some are hotter, others cooler.

If we look to the right or the left of the Milky Way the visible stars are fewer; ***the Milky Way is a high concentration of stars in a narrow zone.*** How can we explain this? Let us think of the shape of the Milky Way system. It is a flattish, disk-shaped aggregation of a hundred bil-

Fig. 36-2. The sun is about midway between the center and the rim of the Milky Way system.



lion stars, a gigantic wheel with the sun near the central plane and about midway between the center and rim (see Fig. 36-2). Actually the sun is about 26,000 light-years from the center, and the diameter of the wheel is 100,000 light-years. We are, of course, inside the system and so we cannot see it as a whole. If we look along the central plane we see a vast concentration of stars; if we look at right angles to the central plane the visible stars are far fewer.

Does the Milky Way move? Actually all galaxies must rotate to maintain their stability. The center of our galaxy is the constellation Sagittarius, and the Milky Way system revolves about this star group. The sun's velocity in the galaxy is about 140 miles per second, but the galaxy is so vast that it takes 200 million years for the sun to make a single revolution.

Is the Milky Way system crowded with stars? A photograph of the Milky Way shows a vast number of stars closely packed together. But this is an illusion. They are so far away that even stars that seem to be almost touching on a photograph must be billions of miles apart. The total volume of all the stars is only a tiny fraction of the space occupied by the galaxy and, even if the number of stars were increased a million fold, the galaxy would not be crowded. Indeed, the British astronomer Jeans once said, "If there were only three wasps over the whole continent of Europe, Europe would be more crowded with wasps than space is with stars."

CONSTELLATIONS

You may have noticed that the brighter stars in the sky are in groups which form patterns. These star patterns are called ***constellations*** and they enable us to divide the sky into regions. Actually there are 90 constellations.

Greek and Arab astronomers, four thousand years ago, associated these patterns with real or fanciful creatures and gave them names commemorating people in their myths and legends. Many of

Fig. 36-3. The best-known constellation is Ursa Major (the Great Bear), often called the Big Dipper. The two stars on the far side of the Dipper point directly to Polaris, the North Star.

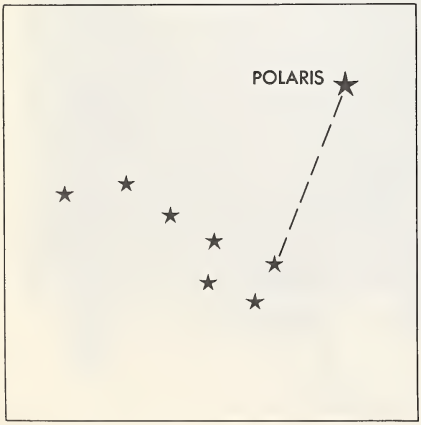
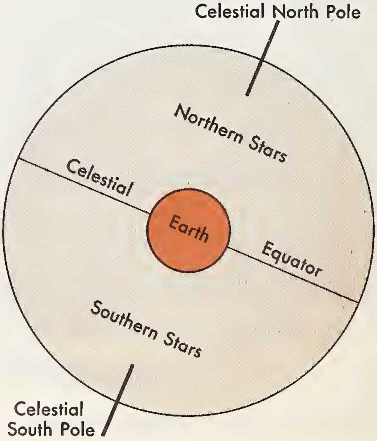


Fig. 36-4. Because the earth rotates on its axis, the stars seem to revolve about the celestial axis. Polaris is the only star in the northern sky which always appears to remain in the same position.



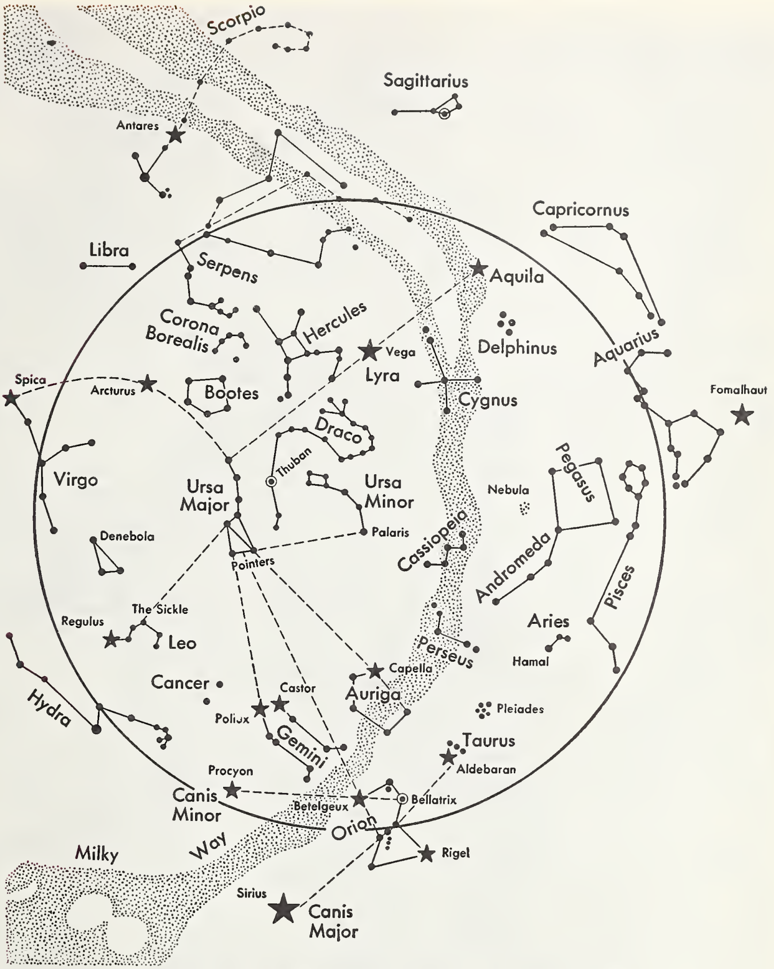


Fig. 36-5. This star map shows the constellations visible in Canada.

these names Leo (Lion), Ursa Major (Great Bear), Taurus (Bull), Draco (Dragon), Scorpio (Scorpion), and Orion (Hunter), are still used today.

Because of the rotation of the earth on its axis, all constellations appear to rotate about Polaris, the pole star. Polaris is almost exactly over the north pole of the earth, so that the earth's axis is directed almost straight at Polaris.

As the earth speeds along its course around the sun, different regions of the heavens are seen from the darkened side of our globe. Thus, some constellations are visible at one season and not at another. Orion, for instance, is easily visible in the winter sky but not during the short summer nights. Hercules, on the other hand, can be seen during the summer but not during the winter.

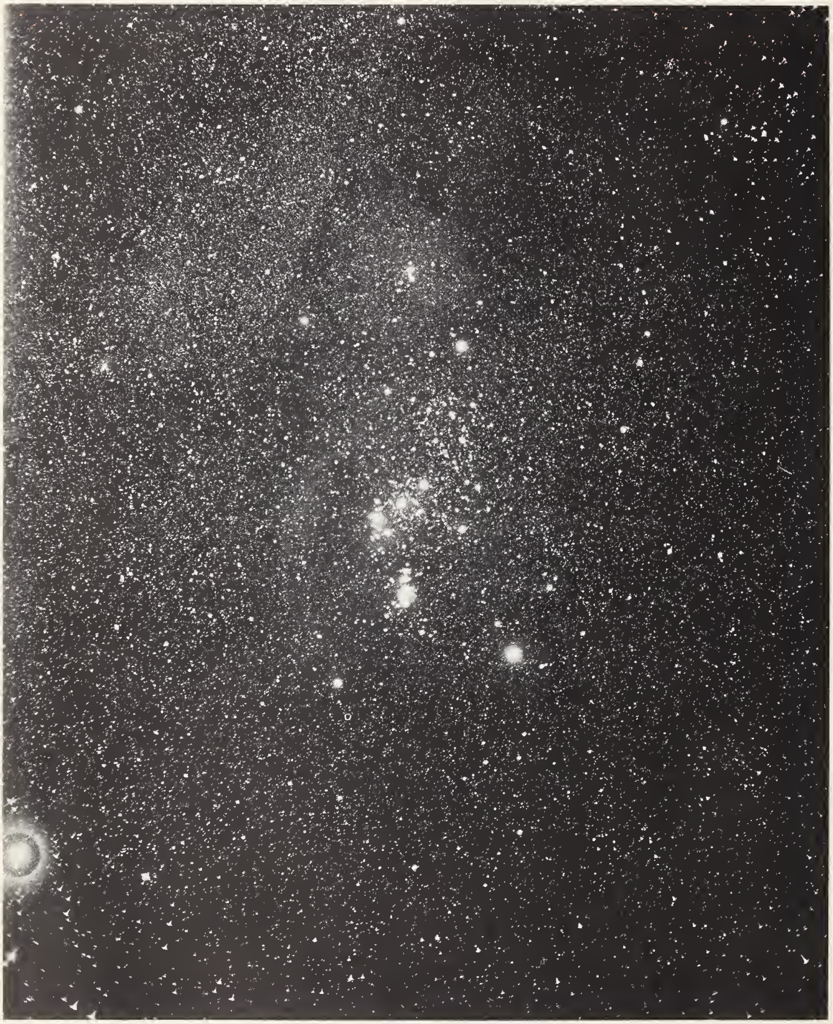


Fig. 36-6. The constellation of Orion. Notice the "belt"—three stars in a row. The two bright stars which are in line with the middle star of the belt are Betelgeuse (above) and Rigel (below). The star in the lower left corner is Sirius, the brightest star in the sky.

Probably the best-known constellation is Ursa Major, the Great Bear. It is often called the Big Dipper because the seven brightest stars are arranged in the shape of a dipper. The two stars forming the far side of the Dipper are called ***the pointers,*** because a line through them across the sky leads to Polaris. Once you have located Ursa Major, you will easily find other nearby constellations with the help of a star map.

A Star Map. To understand a star map we should know what is meant by the "dome of the heavens" with respect to

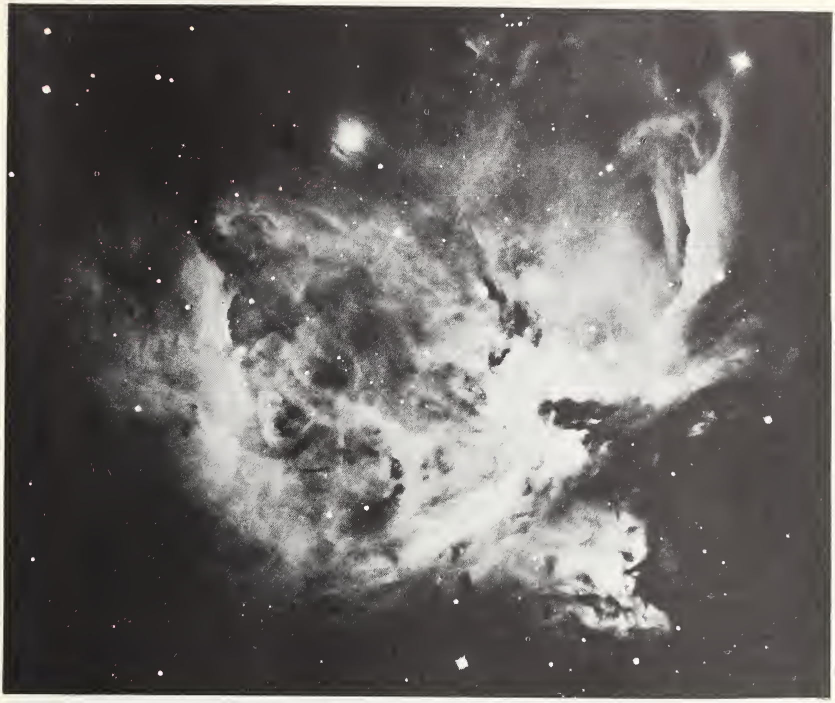


Fig. 36-7. The Great Nebula in Orion. Below Orion's belt are three stars forming the sword. This picture shows the middle "star" in Orion's Sword; it is the brightest of the diffuse nebulae in the sky.

earth. This is shown in Fig. 36-4. Notice that the celestial equator passes through and beyond the earth's equator and that the celestial poles are drawn by projecting a line through the earth's poles (the earth's axis) on to the sky. The Star Map (Fig. 36-5) shows the constellations visible in Canada. The dotted band is the Milky Way. Let us pick out a few constellations on the map. Then look for them in the night sky.

Cassiopeia. The constellation of Cassiopeia has its five main stars arranged like a letter W. It can readily be recognized in the sky in late summer by looking to the northeast. To the south of Cassiopeia you may be able to distinguish a fuzzy patch of light, particularly if you have binoculars. This is the Andromeda Nebula, the nearest galaxy beyond the Milky Way.

Orion. Orion is a magnificent spectacle in the winter sky. Look in a southerly direction and you will see three stars in a row. These constitute the "belt" of Orion. Above the belt is the bright red star Betelgeuse, and below it the bluish star Rigel. The line joining Betelgeuse and Rigel makes a cross with the belt. Below the belt there are three stars forming Orion's sword. The middle star of the sword is fuzzy or hazy. A photograph of this region shows that the hazy patch is actually a huge mass of glowing gas. It is called the Great Nebula of Orion and, seen through a telescope, is a magnificent spectacle. (See Fig. 36-7.) To the left of Orion, and a little below, is Sirius.

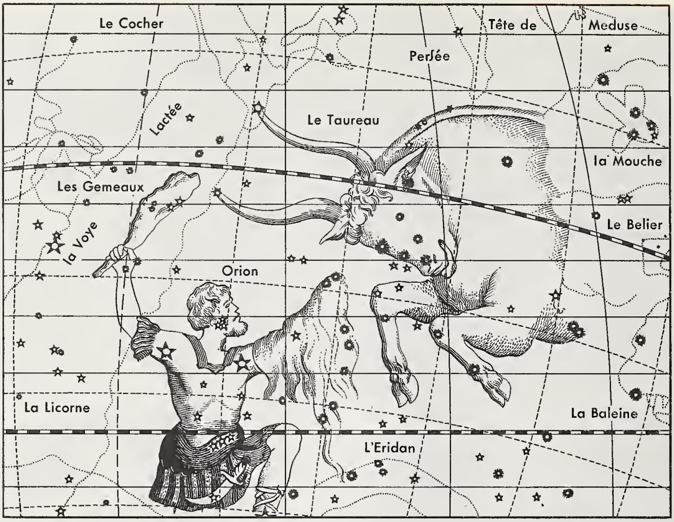


Fig. 36-8. The ancients imagined that Orion (the hunter) chased Taurus (the bull) across the sky.

Sirius is the brightest star in the sky and one of the nearest stars to the earth.

Taurus. According to ancient legend, Orion the hunter is facing Taurus the Bull with club in outstretched arm (See Fig. 36-8). Taurus is the Constellation next to Orion; its brightest star is Aldebaran.

Andromeda. Observation by the unaided eye in the region of Andromeda shows a distinct spot of faint light. With a pair of binoculars we can see an oval area of light. But when examined through a powerful telescope, or on a photographic plate, the full beauty of the oval patch is revealed. This is the great nebula in Andromeda; it is an outside galaxy, our "neighbor galaxy."

The 200-inch Hale reflector has resolved the nebula of Andromeda into stars. In dimensions Andromeda is about the same size as our own galaxy, and the number of its stars is about the same. It is two million light-years away. In other words, two million light-years is the limit of vision with the unaided eye.

It is important to realize that when we see Andromeda we are not only looking into space, but also back into time. The light that enters our eye has been traveling through space for two million years; it left Andromeda two million years ago. Does Andromeda exist today? No one knows for sure. Indeed, the question cannot be definitely answered for another two million years. Why not?

To emphasize again the question of stellar distances, let us imagine we can set off from the earth in a rocket ship capable of traveling with the speed of light. It will take eight minutes to reach the sun, one hour to reach Jupiter and five hours to reach the outermost planet Pluto. Leaving the solar system we head for the nearest star, Alpha Centauri. The trip to Alpha Centauri takes 4½ years, to Sirius 7 years, to Aldebaran 55 years, and to the bright star Rigel 543 years. But we are still in the Milky Way system. To get out by the shortest path may take at least 5000 years. Then, still traveling at the speed of light, we will steer towards

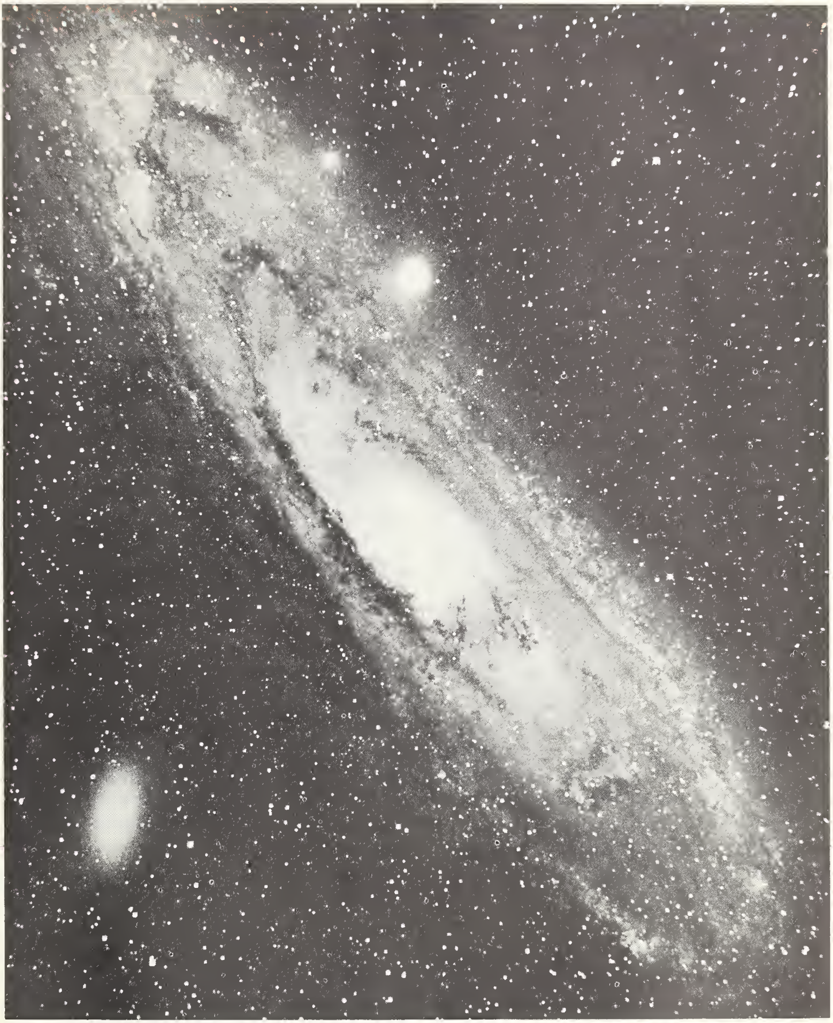


Fig. 36-9. The Great Nebula in Andromeda. It is the nearest galaxy to the Milky Way, two-million light years away.

Andromeda, the nearest galaxy. This part of the journey will take two millions years and, at the end of it, there is a remote chance that we may not find Andromeda —it may have ceased to exist!

THE UNIVERSE

Modern telescopes see far beyond the Milky Way, and photographs made by them reveal vast numbers of galaxies outside our own. Indeed, the 200-inch Hale telescope on Palomar can see a distance of two billion light-years and it can see almost a billion galaxies. Since we can look 2 billion light-years into space in all directions the region we observe is a vast sphere, 4 billion light-years in diameter. And within this sphere 2 billion galaxies are scattered. The Observable Region of the universe could be represented by a small model, a sphere 10 miles in diameter. The galaxies would then be the size of soft balls about 40 yards apart scattered throughout the sphere. On this scale the earth would be too small to be seen, even through a microscope.

As we stated in the first chapter of the book, these galaxies appear to be fleeing away from each other. By measuring the speeds of the galaxies, and their distances away, it is possible to compute the time when the motion began; that is, when all galaxies were compressed together. This was about six billion years ago.

Some scientists believe that the universe was in a rarefied state for almost an eternity of time. Through countless ages it steadily contracted, becoming more and more compressed. Eventually the matter was so compressed that electrons were torn from their orbits and nuclei were packed closely together. Matter was then compressed as much as it could be—its density was millions of times the density of water. Then came the mighty explosion, six billion years ago. This was the birth of the nebular universe as we now see it.

The Birth and Death of Stars. Do stars shine forever or does their fuel supply eventually run out? This question was first raised by the American scientist Hans Bethe in 1938, and it has been studied by astrophysicists ever since. You will recall that Bethe showed that the heat of stars is due to a continuing nuclear reaction in which four hydrogen atoms are converted to one helium nucleus.

How do astronomers find the age of a star? If an engineer knows the rate at which coal burns in his furnace, and the amount of ash, he can find out when the fire was started. Astronomers are faced with a similar problem. If they know the rate at which hydrogen is converted to helium and the amount of helium "ash," they can compute the age of a star. Moreover, if an engineer knows the amount of coal he has and its rate of burning, he can compute the time it will burn. Similarly from the mass of a star, astronomers can compute the amount of hydrogen it contains, and hence how long it will burn. From this information the future life of the star can be calculated. They have found that some stars are old, others are young. The oldest date back to the beginning of the universe; the youngest are not more than a few million years old.

A star is said to be in a stable condition when its force of gravity, pulling material towards its center, is balanced by the gas pressure in its interior, pushing the material outwards. The larger the size of the star, the greater is its gravitational force and therefore the higher must be the interior temperature to overcome this extra force. When a star is stable it neither expands nor contracts. However, the stability of a star is affected by the percentage of helium it contains. Helium is the "ash" of the nuclear reaction formed deep in the interior of stars. As hydrogen changes to helium the internal gas pressure changes and, to offset this change, the star grows bigger and it consumes fuel (hydrogen) more rapidly. As a result, the star steadily increases in brightness and continues to do so until 12 per cent of the hydrogen is used up. At this point, the star can no longer compensate for its high helium content by a slow increase in size and "burning" rate. Instead, it grows rapidly in size and its rate of burning increases enormously. This is a sign that the star is near the end of its active life; within a few million years the ash puts out the "nuclear fires," the star cools, contracts and "dies."

Death of the Sun. Astronomers tell us that the sun has already used up 6 per cent of its hydrogen. Since it consumes about one per cent of hydrogen every billion years, the sun is about 6 billion years old. It also has about 6 billion years to run; it has already lived one-half of its life. In another six billion years the sun will have reached the significant 12 per cent mark; it will have reached old age. Then it will begin to increase in size rapidly until its diameter has increased about 30 times. Its rate of nuclear reaction will also speed up enormously.

How will all this affect the earth? Here we are in the realm of speculation. Most astrophysicists portend catastrophic events. They say the temperature of the earth will rise above the melting point of aluminum (660° C). But long before this happens the seas will have boiled away and all life on the planet will have been wiped out. Nor is this all. As the sun consumes its fuel, it will become cooler, diminish in brightness, contract, and eventually become a white dwarf. On earth, the temperature will fall, the ocean clouds will condense to reform the seas, and ultimately, the seas will freeze to a solid block of ice. This is the predicted end of the earth.

This description is a bleak and grim prospect, even though the end is far off. But it is a part of the process of stellar evolution. Some stars sink into obscurity; others are born. In some ways we in the solar system are fortunate. Our sun is aging slowly—we have yet 6 billion years to live. Will the passing of life on earth mean the end of life in the universe? As stated in the first chapter of the book, most astronomers believe there are many solar systems like ours, and it is therefore probable that there are other earths that support life. But these other earths will pass through the same evolutionary cycle, and they too will come to the same inevitable end.

Things to Remember

A ***light-year*** is the distance light travels in one year.
A ***galaxy*** is a system of billions of stars.
The ***Milky Way*** is the galaxy that includes the sun among its stars.
Constellations are star patterns; they divide the sky into sections.
The best-known constellation is ***Ursa Major,*** the Great Bear.
The ***universe*** consists of billions of galaxies.

Questions

GROUP A

1. State the meaning of the words (*a*) nebula, (*b*) galaxy.
2. The nearest star is 4.5 light-years from the sun. What does this mean?
3. Where is the sun in the Milky Way system?
4. What is a constellation? Give three examples.
5. How would you recognize Cassiopeia in the night sky?
6. What are the ***pointers?***

GROUP B

7. There were three separate steps in man's exploration of space. Explain what they are.
8. What is a galaxy? Illustrate your answer by reference to the Milky Way.
9. Describe the constellation of Orion.

10. "When we look into space, we also look back into time." Explain the meaning of this statement by referring to the great nebula in Andromeda.

11. What is the astronomer's view of the universe?

Things to Do

To study the principal constellations · We know that some of the stars appear to be in groups known as constellations. A study of the constellations is an interesting project. Provide yourself with a star map or a star book, a flashlight whose light has been dimmed by covering the lens with dark red paper, a drawing pad, and pencils. Your star study can begin at the North Star and at the pointers in the Great Bear constellation. Gradually build your knowledge of the constellations by searching the sky for these groups of stars. When you have made a complete observation of the sky, try to draw some star maps of your own. Indicate by dots, larger or smaller, the relative brightness of the stars. By dotted lines you can join together stars belonging to the same constellation.

GLOSSARY

Acoustics. Sound-absorbing and sound-reflecting properties of an auditorium.

Aeration. Oxidation of organic impurities in raw water by impregnating it with common air.

Air mass. A mass of cold, dry air from the polar regions or of warm, moist air from the semi-tropics.

Alloy. A solid mixture of metals. The alloying metal is added to give a desired quality, such as durability.

Alpha particles. Positively charged particles (nuclei of helium atoms) ejected in some radioactive disintegrations.

Alternating current (A. C.). A current that flows first in one direction and then reverses and flows in the opposite direction.

Alternator. A generator for producing alternating current.

Ammeter. An instrument for measuring an electric current in a circuit.

Amplifier. A vacuum tube device to increase the voltage in a circuit.

Amplitude. Extent of a vibratory movement measured from the mean position to the extreme.

Amplitude modulation (AM). Varying of the intensity of radio waves according to the sound variations to be broadcast.

Anemometer. An instrument that measures wind velocity.

Aneroid. Without liquid. An aneroid barometer.

Anneal. To make glass less brittle by heating it until it is soft and then cooling it.

Antibiotics. Chemicals produced by living organisms which destroy harmful bacteria.

Anticline. An upfold or arch of stratified rock.

Anticyclone. A large whirlpool of air at high pressure moving clockwise.

Antiseptics. Chemicals that prevent infection of wounds by bacteria.

Archeozoic era. Period of earliest life in geologic history.

Armature. Coils of insulated wire supported on a shaft that revolves in a magnetic field, used to induce an electric current.

Atomic weight. Sum of neutrons and protons in nucleus of an atom.

Audio amplifier. A device to increase the power of a radio signal.

Axis of earth. An imaginary line about which the earth revolves.

Bacteria. Microscopic, one-celled, parasitic plants.

Ballistic missile. A giant rocket which may be guided or unguided.

Batholith. A great mass of intruded igneous rock originating deep in the earth.

Beaufort scale. Scale from 0 to 12 for measuring wind velocity.

Beriberi. A disease, causing general debility, resulting from a lack of vitamin B in the diet.

Beta particles. Negatively charged particles (electrons) emitted by certain radioactive substances.

Block mountains. Mountains formed by massive blocks due to faulting.

British thermal unit. (B.t.u.). The amount of heat required to raise the temperature of one pound of water one degree Fahrenheit.

Carbochemicals. Chemicals derived from the distillation of coal.

Carbohydrates. A group of compounds composed of carbon, hydrogen, and oxygen. Sugars and starches are carbohydrates.

Carboniferous period. The geologic period in the late Paleozoic. Coal was formed from the vegetation of this period.

Catalyst. A substance that speeds up a reaction but is not itself affected chemically.

Cathode. The negative pole of an electrolytic cell.

Cathode rays. A stream of electrons.

Cenozoic era. The era of recent life in geologic history.

Chain reaction. A reaction which causes further reactions of the same kind, therefore self-sustaining.

Circuit. A complete path of an electric current.

Cirrus cloud. A high, thin cloud consisting of ice crystals.

Coarse-grained rock. An igneous rock with large crystals; for example, granite.

Cold front. The face of a cold air mass that pushes back a warm air mass.

Comet. A heavenly body, having a head and a long tail, which travels around the sun.

Commutator. A device for reversing the direction of an electric current.

Continental glacier. An ice sheet covering a large area of land.

Continental shelf. A submarine plane bordering nearly every continent.

Constellation. A group of stars that seem to form a figure or pattern.

Contour plowing. A method of soil conservation in which furrows are plowed to follow the curvature of a hill rather than up and down the hill.

Cosmic rays. Extremely high-penetrating rays formed by transmutation of atoms in space.

Coulomb. The quantity of electricity conveyed by one ampere in one second.

Cracking of petroleum. The process of breaking up kerosene molecules to obtain gasoline.

Cumulus clouds. Large, fluffy clouds that develop into thunderstorms.

Cyclone. A whirlpool of air at low pressure moving counterclockwise.

Decibel. Unit for measuring the loudness of sounds.

Declination. The angle which a magnetic needle makes with the geographic meridian.

Destructive distillation. Decomposition of a complex substance by heat and the condensation of some of the vapors.

Detergent. A cleansing agent, usually a soapless soap.

Deuterium. A heavy isotope of hydrogen.

Deuteron. A nucleus of deuterium.

Dew point. The temperature to which the atmosphere must be cooled to produce condensation.

Diatoms. Microscopic plants which are the source of oil deposits on the sea floor.

Dielectric. A non-conducting material.

Digestion. Breaking food down chemically by the action of enzymes.

Diode. A vacuum tube used to rectify alternating current to direct current.

Direct current (D. C.). A current that always flows in the same direction.

Doldrums. The region on the earth between 10° N. and 10° S. in which there is little or no wind.

Domain magnetic theory. A theory that explains the magnetization of iron.

Drainage basin. An area of land drained by a river system.

Dust-cloud theory. One of the theories that explains the origin of the solar system.

Edison effect. The "boiling out" of electrons from a hot filament.

Electromagnet. An iron rod magnetized by an electric current flowing through the coil of wires wound around it.

Electromotive force. The force that causes an electric current to flow.

Emulsion. A dispersion of globules of one liquid in another liquid.

Equinox. The time when day and night are everywhere of equal length.

Erosion. The wearing away of rock surfaces and removal of loosened material.

Estuary. An arm of the sea at the lower end of a river.

Evolution. The process by which organisms pass through a series of changes in response to changing conditions.

Exciter. A battery which supplies a current to produce a magnetic field in a generator or motor.

Fall-out. Radioactive dust carried into the upper atmosphere following an atomic bomb explosion, and later brought down by rain.

Fault. In geology, a fracture on the earth in which layers of rock slide up or down along the break.

Filament. Thread-like metal conductor in a light bulb.

Filtrate. A filtered solution.

Fluorescence. The emission of white light by a chemical, such as zinc sulfide, absorbing radiations from mercury vapor.

Fossils. Bones, shells, or impressions in rocks of animals or plants that lived in past ages.

Frequency. The number of vibrations (or cycles) per second.

Frequency Modulation (FM). Varying of the frequency of radio waves according to the sound variations to be broadcast.

Front. Boundary between a cold-air and a warm-air mass.

Fundamental. The component of lowest frequency or greatest wave length.

Fuse. An alloy with a low melting point used in an electric circuit to break the circuit when an excessive current flows through it.

Fusion. A combination of the nuclei of light elements with a release of energy; for example, the fusion of hydrogen to helium.

Galaxy. A concentration of millions of stars, such as the Milky Way.

Gamma rays. Rays given off by radioactive substances; they have no electric charge.

Generator. A machine that converts mechanical energy into electrical energy.

Geosyncline. An enormous subterranean trough in the earth's crust in which deposits collect.

Great circle. The equator, or a meridian of longitude.

Gravitation. The force by which every body in the universe is attracted to every other body.

Half-life. The time required for half the atoms of a radioactive element to disintegrate.

Hard water. Water that does not lather with soap because of calcium and magnesium salts in solution.

High. An anticyclone; a whirlpool of air at high pressure, generally bringing fair weather.

High-frequency current. A rapid vibration of charges across a spark gap.

Hurricane. An atmospheric disturbance of low pressure that originates over water near the equator and is accompanied by violent winds.

Hydraulic brakes. Brakes which depend on the transmission of force through a fluid.

Hygrometer. An instrument for measuring moisture in the air.

Ice Ages. The periods during the Cenozoic era when ice covered a large part of the continent of North America.

Impervious rock. Rock through which water cannot penetrate.

Induced current. Production of an electric current by moving a conductor in a magnetic field.

Invertebrates. Animals without backbones.

International date line. Meridian 180°. Vessels crossing the line going west set the date forward by one day, and vice versa.

Ionosphere. Layer of air high above the earth which conducts electricity and transmits radio waves over long distances.

Isobars. Lines on a weather map connecting points having the same barometric pressure.

Isotopes. Atoms of the same element with different atomic weights due to a different number of neutrons in the nucleus.

Jet stream. A high-speed stream of air in the troposphere. It blows from west to east at an altitude of about 30,000 feet.

Kilowatt-hour. Unit for measuring electric energy; 1.34 horsepower-hour.

Latitude and longitude. Latitude is the number of degrees north or south of the equator; longitude, the number of degrees east or west of Greenwich, England.

Light-year. The unit of measurement for the distances of stars from the earth; the distance that light travels in a year.

Lodestone. Iron ore, magnetite, with magnetic properties. Used in earliest compasses.

Longitudinal waves. Sound waves. Waves in which air molecules move in the same line as the waves.

Low. A cyclone; a whirlpool of air at low pressure, generally bringing stormy weather.

Magma. Molten rock within the earth, which ultimately cools to form igneous rock.

Magnetic induction. A magnetizable body becomes magnetized in a magnetic field.

Maritime climate. Small range between summer and winter temperatures.

Mesozoic era. In the earth's history, the period of middle life.

Metamorphic rock. A rock that has been changed by chemical action and pressure.

Meteor. A piece of rock or dust that enters the earth's atmosphere at great speed and becomes incandescent because of the resistance of the air.

Meteorite. A meteor that reaches the earth.

Meteorology. The branch of science dealing with the weather.

Millibar. A unit of atmospheric pressure. Pressure in millibars is shown on weather maps.

Milky Way system. A galaxy of millions of stars which includes the sun.

Monsoon. A seasonal prevailing wind that blows from one direction part of the year and from the opposite direction the rest of the year.

Moraine. Stones, boulders, clay, and other matter deposited by melting glaciers.

Narcotic. A drug that relieves pain but induces stupor.

Neap tide. The lowest tide in a lunar month.

Nimbus cloud. A rain cloud, uniformly gray and extending over the entire sky.

Nucleus of an atom. The core. Consists of protons and neutrons. Always positively charged.

Occluded front. A meteorological situation in which a warm-air mass is wedged off the ground between two cold-air masses.

Orbit of earth. Path described by earth around the sun.

Oscillator tube. Generates alternating currents of specific frequencies in a radio transmitter.

Overtone. A sound with twice the frequency of the fundamental tone.

Paleozoic era. The period of ancient life in earth history.

Parallel circuit. An electric circuit in which like poles of a battery are connected to each other.

Pasteurization. Controlled heating of milk to destroy bacteria.

Petrochemicals. Chemicals obtained from cracking the molecules of petroleum.

Phases of moon. A recurring cycle of changes from new moon, to first quarter, to full moon, to last quarter.

Photoelectric cell. A device for changing light energy into electric energy.

Pitch. The quality of a sound due to the number of vibrations per second.

Planet. One of the nine heavenly bodies that revolve around the sun.

Precipitation. Rain, snow, hail, etc., that descend on the earth from the atmosphere.

Prevailing westerlies. The winds that blow from the southwest between latitudes 30° and 60° of the Northern Hemisphere, and from the northwest in the same latitudes of the Southern Hemisphere.

Proteins. A class of organic compounds which contain nitrogen as well as carbon, hydrogen, and oxygen. An essential part of a balanced diet.

Proterozoic era. Period of primitive life in earth history.

Radar. Radio direction finding and ranging. A device that locates objects through dense fog and clouds by means of radio waves of extremely high frequency.

Radioactive wastes. The waste "ash" from a reactor; highly radioactive.

Radioactivity. Ejection of rays (alpha, beta, and gamma) from the nuclei of radioactive atoms.

Radiotelephony. Telephony by the aid of radio waves; that is, without connective wires.

Radiotelescope. An instrument that picks up radio waves from planets and stars.

Radiosonde. Apparatus carried by a balloon to the stratosphere to measure air conditions.

Reactor. An atomic furnace used for the fission of uranium.

Reciprocal. One (unity) divided by the quantity. The reciprocal of a fraction is the fraction inverted.

Relative humidity. Ratio of the actual amount of vapor present in the atmosphere at a given temperature to the maximum amount of vapor the atmosphere can hold at that temperature.

Respiration. The slow burning, or oxidation, of food in the cells of an animal body.

Resonance. Reinforcement of sound. The response of an air column to a vibrating tuning fork.

Reverberation. Reflection of sound waves. An echo.

Rotor. The rotating part of an electric generator.

Scurvy. A disease characterized by hemorrhage of mucous membranes and gums. Due to lack of vitamin C.

Short circuit. A circuit formed when an electric current flows through a small resistance rather than through the regular circuit of larger resistance.

Sound barrier. Maximum air resistance against a plane traveling at the speed of sound.

Spectrohelioscope. An instrument for observing solar eruptions.

Specific gravity. Ratio of the weight of a substance with the weight of a similar volume of water.

Spring tide. A tide greater than usual. Caused when sun and moon are in conjunction or opposition.

Stalactite and stalagmite. Formations of calcium carbonate resembling icicles on the ceiling and floor of limestone caves.

Storage battery. Converts chemical energy into electrical energy. Can be recharged by a current.

Stator. The stationary (non-revolving) part of a generator.

Stratosphere. The outer layer of atmosphere about 40,000 feet above the earth's surface in which there is no weather.

Sunspots. Cooler places on the surface of the sun that appear as dark patches.

Syncline. A downward fold in sedimentary rocks.

Tape recorder. A device that makes recordings on magnetizable tape.

Teletype. A system of communication in which an instrument similar to a typewriter is used.

Terracing. A method of soil conservation in which a long slope is broken into shorter slopes to prevent soil from being washed away.

Time belt. A belt 15 degrees of longitude wide. Adjacent time belts differ in time by one hour.

Tornado. A small atmospheric disturbance of low pressure, starting on a very hot day and accompanied by winds of high velocity that are highly destructive.

Tradewinds. Prevailing winds in the belts 10°-25° north and south of the equator. In the Northern Hemisphere they blow toward the equator from the northeast, in the Southern Hemisphere from the southeast.

Tranquilizers. Drugs to relieve mental stress.

Transformer. A device that changes the voltage of an alternating current. *Step-up transformer* increases voltage; *step-down transformer* reduces voltage.

Transmitter. Part of telephone into which one speaks. Key of telegraph set. Device that sends out radio waves in broadcasting.

Transverse waves. Motion of particles of medium in a plane perpendicular to direction of waves.

Transistor. A small amplifying device, used in place of vacuum tube.

Triode. A vacuum tube with a grid that controls current and voltage; it is used in amplifiers.

Tritium. The heaviest isotope of hydrogen.

Troposphere. The thin shell of air surrounding the earth in which weather changes take place.

Unconformity. Surface of separation between newer and older rock strata.

Ultrasonics. Sounds with frequencies higher than can be heard by the human ear.

Vacuum tube. A sealed tube with most of air removed so that electric discharges can pass between metal electrodes which project into the tube.

Vapor lamp. Light of lamp is a glowing gas caused by passing an electric current through the gas.

Vertebrates. Animals with backbones.

Vitamins. Constituents of foods in their natural state, very small quantities of which are essential for health and nutrition.

Warm front. Part of front between warm and cold-air masses, and warm air replaces cold air at ground level.

Water cycle. Repeated evaporation and condensation in the atmosphere.

Water table. The level of underground water.

Wave length. Distance between crests of successive waves.

Weathering. The wearing away of rocks by wind, rain, cold, and heat.

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MODERN CLASSIFICATION OF THE

TRANSITION

Period	Group I A		II A		III B		IV B		V B		VI B		VII B			
1																
2	2 1	3 Li 6.940	2 2	4 Be 9.013												
3	2 8 1	11 Na 22.991	2 8 2	12 Mg 24.32												
4	2 8 8 1	19 K 39.100	2 8 8 2	20 Ca 40.08	2 8 9 2	21 Sc 44.96	2 8 10 2	22 Ti 47.90	2 8 11 2	23 V 50.95	2 8 13 1	24 Cr 52.01	2 8 13 2	25 Mn 54.94	2 8 14 2	26 Fe 55.85
5	2 8 18 8 1	37 Rb 85.48	2 8 18 8 2	38 Sr 87.63	2 8 18 9 2	39 Y 88.92	2 8 18 10 2	40 Zr 91.22	2 8 18 12 1	41 Nb 92.91	2 8 18 13 1	42 Mo 95.95	2 8 18 13 2	43 Tc (99)	2 8 18 15 1	44 Ru 101.1
6	2 8 18 18 8 1	55 Cs 132.91	2 8 18 18 8 2	56 Ba 137.36	57-71 See Lanthanide Series		2 8 18 32 10 2	72 Hf 178.58	2 8 18 32 11 2	73 Ta 180.95	2 8 18 32 12 2	74 W 183.86	2 8 18 32 13 2	75 Re 186.22	2 8 18 32 14 2	76 Os 190.2
7	2 8 18 32 18 8 1	87 Fr (223)	2 8 18 32 18 8 2	88 Ra 226.05	89-100 See Actinide Series											

NOTE "A" groups consist of elements where the outside ring carries the valence electrons.
"B" groups are the transition groups where a shell other than the outermost is being filled.
Vertical column to the left of the element shows ring structure. The heavy step line represents the division between metals and non-metals. The heavy vertical line separates the inert elements.

Lanthanide Series	2 8 18 18 9 2	57 La 138.92	2 8 18 19 9 2	58 Ce 140.13	2 8 18 20 9 2	59 Pr 140.92	2 8 18 22 8 2	60 Nd 144.27	2 8 18 23 8 2	61 Pm (145)	2 8 18 24 8 2	62 Sm 150.35
Actinide Series	2 8 18 32 18 9 2	89 Ac 227	2 8 18 32 19 9 2	90 Th 232.05	2 8 18 32 20 9 2	91 Pa 231	2 8 18 32 21 9 2	92 U 238.07	2 8 18 32 22 9 2	93 Np (237)	2 8 18 32 23 9 2	94 Pu (242)

ELEMENTS SHOWING ELECTRON STRUCTURES

ELEMENTS

	VIII B				I B		II B		III A		IV A		V A		VI A		VII A		Group O
1	1 **H** 1.008																	2	2 **He** 4.003
								2 3	5 **B** 10.82	2 4	6 **C** 12.011	2 5	7 **N** 14.008	2 6	8 **O** 16.000	2 7	9 **F** 19.00	2 8	10 **Ne** 20.183
								2 8 3	13 **Al** 26.98	2 8 4	14 **Si** 28.09	2 8 5	15 **P** 30.975	2 8 6	16 **S** 32.066	2 8 7	17 **Cl** 35.457	2 8 8	18 **A** 39.944
2 8 15 2	27 **Co** 58.94	2 8 16 2	28 **Ni** 58.71	2 8 18 1	29 **Cu** 63.54	2 8 18 2	30 **Zn** 65.38	2 8 18 3	31 **Ga** 69.72	2 8 18 4	32 **Ge** 72.60	2 8 18 5	33 **As** 74.91	2 8 18 6	34 **Se** 78.96	2 8 18 7	35 **Br** 79.916	2 8 18 8	36 **Kr** 83.8
2 8 18 16 1	45 **Rh** 102.91	2 8 18 18	46 **Pd** 106.7	2 8 18 18 1	47 **Ag** 107.880	2 8 18 18 2	48 **Cd** 112.41	2 8 18 18 3	49 **In** 114.82	2 8 18 18 4	50 **Sn** 118.70	2 8 18 18 5	51 **Sb** 121.76	2 8 18 18 6	52 **Te** 127.61	2 8 18 18 7	53 **I** 126.91	2 8 18 18 8	54 **Xe** 131.30
2 8 18 32 17	77 **Ir** 192.2	2 8 18 32 17 1	78 **Pt** 195.09	2 8 18 32 18 1	79 **Au** 197.0	2 8 18 32 18 2	80 **Hg** 200.61	2 8 18 32 18 3	81 **Tl** 204.39	2 8 18 32 18 4	82 **Pb** 207.21	2 8 18 32 18 5	83 **Bi** 209.00	2 8 18 32 18 6	84 **Po** 210	2 8 18 32 18 7	85 **At** (211)	2 8 18 32 18 8	86 **Rn** 222

Atomic weight in brackets denotes the mass number of the isotope of longest known half-life.

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2 8 18 25 8 2	63 **Eu** 152.0	2 8 18 25 9 2	64 **Gd** 157.26	2 8 18 26 9 2	65 **Tb** 158.93	2 8 18 28 8 2	66 **Dy** 162.51	2 8 18 29 8 2	67 **Ho** 164.94	2 8 18 30 8 2	68 **Er** 167.27	2 8 18 31 8 2	69 **Tm** 168.94	2 8 18 32 8 2	70 **Yb** 173.04	2 8 18 32 9 2	71 **Lu** 174.99
2 8 18 32 24 9 2	95 **Am** (243)	2 8 18 32 25 9 2	96 **Cm** (245)	2 8 18 32 26 9 2	97 **Bk** (245)	2 8 18 32 27 9 2	98 **Cf** (248)	2 8 18 32 28 9 2	99 **E** (255)	2 8 18 32 29 9 2	100 **Fm** (252)		101 **Mv** (256)		102 **No** (253)		

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